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Airbus A380

2008/9 Schools Wikipedia Selection. Related subjects: Air & Sea transport

The **Airbus A380** is a double-deck, four-engine airliner manufactured by the European corporation Airbus, an EADS subsidiary. The largest passenger airliner in the world, the A380 made its maiden flight on 27 April 2005 from Toulouse, France, and made its first commercial flight on 25 October 2007 from Singapore to Sydney with Singapore Airlines. The aircraft was known as the **Airbus A3XX** during much of its development phase, but the nickname **Superjumbo** has since become associated with it.

The A380's upper deck extends along the entire length of the fuselage. This allows for a cabin with 50% more floor space than the next largest airliner, the Boeing 747-400, and provides seating for 525 people in standard three-class configuration or up to 853 people in full economy class configuration. The A380 is offered in passenger and freighter versions. The A380-800, the passenger model, is the largest passenger airliner in the world, superseding the Boeing 747, but has a shorter fuselage than the Airbus A340-600 which is Airbus' next biggest passenger aeroplane. The A380-800F, the freighter model, is offered as one of the largest freight aircraft, with a listed payload capacity exceeded only by the Antonov An-225. The A380-800 has a design range of 15,200 kilometres (8,200 nmi), sufficient to fly from New York to Hong Kong for example, and a cruising speed of Mach 0.85 (about 900 km/h or 560 mph at cruise altitude).

Airbus A380

A300 · A310 · A320 · A330 · A340 · A350 · **A380**

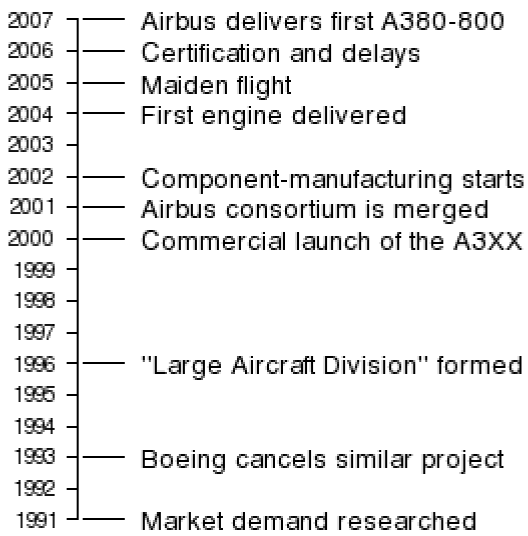
The A380 during its World Tour flight 2006-2007.

Type	Airliner
Manufacturer	Airbus
Maiden flight	27 April 2005
Introduced	25 October 2007 with Singapore Airlines
Primary user	Singapore Airlines
Produced	2002 – present
Number built	11 as of December 2007
Program cost	€12 ($17.1) billion
Unit cost	$319.2 million

History

Development

Airbus started the development of a very large airliner (termed Megaliner by Airbus in the early development stages) in the early 1990s, both to complete its own range of products and to break the dominance that Boeing had enjoyed in this market segment since the early 1970s with its 747. McDonnell Douglas pursued a similar strategy with its ultimately unsuccessful MD-12 design. As each manufacturer looked to build a successor to the 747, they knew there was room for only one new aircraft to be profitable in the 600 to 800 seat market segment. Each knew the risk of splitting such a niche market, as had been demonstrated by the simultaneous debut of the Lockheed L-1011 and the McDonnell Douglas DC-10: both planes met the market’s needs, but the market could profitably sustain only one model, eventually resulting in Lockheed's departure from the civil airliner business. In January 1993, Boeing and several companies in the Airbus consortium started a joint feasibility study of an aircraft known as the Very Large Commercial Transport (VLCT), aiming to form a partnership to share the limited market.




The first completed A380 at the "A380 Reveal" event in Toulouse.

In June 1994, Airbus began developing its own very large airliner, designated the A3XX. Airbus considered several designs, including an odd side-by-side combination of two fuselages from the A340, which was Airbus’s largest jet at the time. The A3XX was pitted against the VLCT study and Boeing’s own New Large Aircraft successor to the 747, which evolved into the 747X, a stretched version of the 747 with the fore body "hump" extended rearwards to accommodate more passengers. The joint VLCT effort ended in July 1996, and Boeing suspended the 747X program in January 1997. From 1997 to 2000, as the East Asian financial crisis darkened the market outlook, Airbus refined its design, targeting a 15 to 20 percent reduction in operating costs over the existing Boeing 747-400. The A3XX design converged on a double-decker layout that provided more passenger volume than a traditional single-deck design.


Fatigue Test Airframe MSN5001 at IMA in Dresden 2005

On 19 December 2000, the supervisory board of newly restructured Airbus voted to launch a €8.8 billion program to build the A3XX, re-christened as the A380, with 55 orders from six launch customers. The A380 designation was a break from previous Airbus families, which had progressed sequentially from A300 to A340. It was chosen because the number 8 resembles the double-deck cross section, and is a lucky number in some Asian countries where the aircraft was being marketed. The aircraft’s final configuration was frozen in early 2001, and manufacturing of the first A380 wing box component started on 23 January 2002. The development cost of the A380 had grown to €11 billion when the first aircraft was completed.

Boeing, meanwhile, resurrected the 747X programme several times before finally launching the 747-8 Intercontinental in November 2005 (with entry into service planned for 2009). Boeing chose to develop a derivative for the 400 to 500 seat market, instead of matching the A380's capacity.

Testing

A380 MSN001 about to land after its maiden flight

Five A380s were built for testing and demonstration purposes. The first prototype, serial number MSN001 and registration F-WWOW, was unveiled at a ceremony in Toulouse on 18 January 2005. Its maiden flight took place at 8:29 UTC (10:29 a.m. local time) 27 April 2005. The prototype, equipped with Trent 900 engines, departed runway 32L of Toulouse Blagnac International Airport with a flight crew of six headed by chief test pilot Jacques Rosay, carrying 20 tonnes (22 short tons) of flight test instrumentation and water ballast. The take-off weight of the aircraft was 421 tonnes (464 short tons); although this was only 75 percent of its maximum take-off weight, it was the heaviest take-off weight of any passenger airliner ever flown.

In mid-November 2005, the A380 embarked on a tour of Southeast Asia and Australia for promotional and for long-haul flight testing purposes, visiting Singapore, Brisbane, Sydney, Melbourne and Kuala Lumpur. During this tour, the livery of Singapore Airlines, Qantas and Malaysia Airlines were applied in addition to the Airbus house livery. On 19 November, an A380 flew in full Emirates livery at the Dubai Air Show.

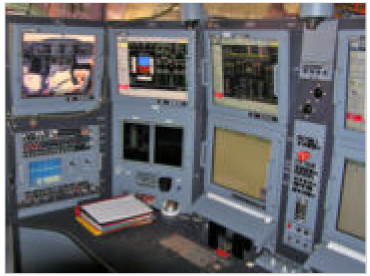
At the 2005 Paris Air Show

On 1 December 2005, the A380 achieved its maximum design speed of Mach 0.96, in a shallow dive, completing the opening of the flight envelope. The aircraft's maximum allowed operational speed is lower, at Mach 0.89, and its cruising speed is Mach 0.85.

On 10 January 2006, the A380 made its first transatlantic flight to Medellín in Colombia, to test engine performance at a high altitude airport. It arrived in North America on 6 February, landing in Iqaluit, Nunavut in Canada for cold-weather testing. The same aircraft then flew to Singapore to participate in the Asian Aerospace 2006 exhibition, in full Singapore Airlines livery.

On 26 March 2006, the A380 underwent evacuation certification in Hamburg in Germany. With 8 of the 16 exits blocked, 853 passengers and 20 crew left the aircraft in 78 seconds, less than the 90 seconds required by certification standards. Three days later, the A380 received European Aviation Safety Agency (EASA) and United States Federal Aviation Administration (FAA) approval to carry up to 853 passengers.

The first A380 planned for delivery to a customer, serial number MSN003 and registration F-WWSA, took to the air in May 2006. The maiden flight of the first A380 with GP7200 engines serial number MSN009 and registration F-WWEA took place on 25 August 2006.



Flight test engineer's station on the lower deck of A380 F-WWOW at the 2006 Farnborough International Airshow.

On 4 September 2006, the first full passenger-carrying flight test took place. The aircraft flew from Toulouse with 474 Airbus employees on board, in the first of a series of flights to test passenger facilities and comfort. In November 2006, a further series of route proving flights took place to demonstrate the aircraft's performance for 150 flight hours under typical airline operating conditions.

Airbus obtained type certificate for the A380-841 and A380-842 model from the EASA and FAA on 12 December 2006 in a joint ceremony at the company's French headquarters. The A380-861 model obtained the type certificate 14 December 2007.

As of December 2007, eleven A380s have flown, and the five A380s in the test programme had logged over 4,565 hours during 1,364 flights, including route proving and demonstration flights. As of December 11th 2007, the A380 has visited 26 countries: Argentina, Australia, Brazil, Canada, China, Colombia, Ethiopia, France, Germany, Iceland, India, Ireland, Japan, Malaysia, Norway, Philippines, Portugal, Singapore, South Africa, South Korea, Spain, Thailand, Turkey, the United Arab Emirates, the United Kingdom, United States of America and Vietnam.

Delivery delays

Initial production of the A380 was plagued by delays attributed to the 530 km (330 miles) of wiring in each aircraft. Airbus cited as underlying causes the complexity of the cabin wiring (100,000 wires and 40,300 connectors), its concurrent design and production, the high degree of customisation for each airline, and failures of configuration management and change control. Deliveries would be pushed back by nearly two years.

Specifically, it would appear that German and Spanish Airbus facilities continued to use CATIA version 4, while British and French sites migrated to version 5. This caused overall configuration management problems, at least in part because wiring harnesses manufactured using aluminium rather than copper conductors necessitated special design rules including non-standard dimensions and bend radii: these were not easily transferred between different versions of the software.

While Airbus attributes the delays entirely to wiring, industry analyst Richard Aboulafia, noting that the first A380 will be around 5.5 tons heavier than intended, speculates that the weight problems "[go] a long way in explaining the delay", and that "wiring alone did not explain what we were all hearing. It sounds like weight-reduction design changes are a big part of the delay, too."

Airbus announced the first delay in June 2005 and notified airlines that delivery would slip by six months, with Singapore Airlines expecting the first A380 in the last quarter of 2006, Qantas getting its first delivery in April 2007 and Emirates receiving aircraft before 2008. This reduced the number of planned deliveries by the end of 2009 from about 120 to 90–100.

On 13 June 2006, Airbus announced a second delay, with the delivery schedule undergoing an additional shift of six to seven months. Although the first delivery was still planned before the end of 2006, deliveries in 2007 would drop to only 9 aircraft, and deliveries by the end of 2009 would be cut to 70–80 aircraft. The announcement caused a 26% drop in the share price of Airbus's parent, EADS, and led to the departure of EADS CEO Noël Forgeard, Airbus CEO Gustav Humbert, and A380 programme manager Charles Champion. In the wake of the new delay, Malaysia Airlines and ILFC were reported to be considering the cancellation of their orders. Launch customers Singapore Airlines, Emirates and Qantas also were reported to be angered by the delays and expecting compensation. However, on 21 July 2006, Singapore Airlines ordered a further 9 A380s and stated that Airbus had *"demonstrated to our satisfaction that the engineering design for the A380 is sound [and that] it has performed well in flight and certification tests and the delays in its delivery have been caused more by production, rather than technical, issues."*

On 3 October 2006, upon completion of a review of the A380 program, the then CEO of Airbus, Christian Streiff, announced a third delay, pushing the first delivery for Singapore Airlines to October 2007, to be followed by 13 deliveries in 2008, 25 in 2009, and the full production rate of 45 aircraft per year in 2010. The delay also increased the earnings shortfall projected by Airbus through 2010 to €4.8 billion. The customer with the largest A380 order, Emirates, saw its first delivery pushed back to August 2008 and said as a result that it was considering scaling back its order, potentially in favour of the rival Boeing 747-8. However, Emirates never scaled back the order but placed additional orders for A380s in 2007. Virgin Atlantic deferred its deliveries by four years, to 2013. The third delay was followed by the first cancellations to hit the A380 programme. On 7 November 2006 FedEx cancelled its order for 10 A380F freighters in favour of 15 Boeing 777 Freighters. In March 2007, the last remaining customer for the A380F, UPS, announced the cancellation of its order. Airbus suspended work on the freighter version in order to concentrate on delivering the passenger version, but said the freighter remained on offer. As of March 2007, Airbus estimated a 2014 entry into service for the A380F.

Entry into service

The first aircraft sold, MSN003, was handed over on 15 October 2007, following a lengthy acceptance test phase, and entered into service on 25 October 2007 with a commercial flight between Singapore and Sydney (flight number SQ380). The plane was given the registration number 9V-SKA. Two months later Singapore Airlines CEO Chew Choong Seng said that the A380 was performing better than both the airline and Airbus had anticipated, burning 20% less fuel per passenger than the airline's existing 747-400 fleet. Singapore Airlines plans to use its first three aircraft, in a 471-seat configuration, on its London–Singapore–Sydney service; until then the A380 will be used between Sydney and Singapore. Subsequent routes may include the Singapore–San Francisco route via Hong Kong, as well as direct flights to Paris, Narita and Frankfurt. The second A380 for Singapore Airlines, MSN005, was handed over by Airbus on 11 January 2008 and registered as 9V-SKB.

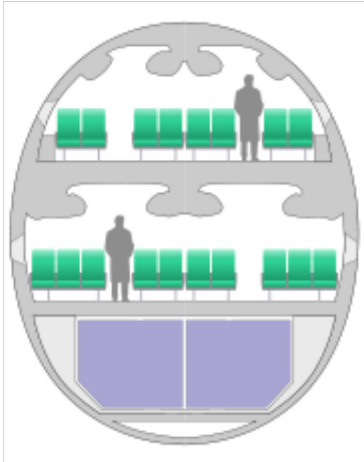

Singapore Airlines is, as of 12 January 2008, the sole A380 operator with the plane in commercial service.

The first aircraft for Qantas (second airline to take delivery of the A380), MSN014, was approaching final wiring installation in September 2007 and was expected to be shipped to Hamburg for cabin fitting out by the end of 2007. Qantas has announced it will use the A380, in a 450-seat configuration, on its Melbourne and Sydney to Los Angeles and Melbourne and Sydney to London routes.

The first Engine Alliance powered A380 MSN011, which is due to enter service with Emirates Airline, had its maiden flight on 4 September 2007. Emirates will receive the aircraft in September 2008 and will initially deploy the plane on its Australian services to Sydney and shortly after to Melbourne. Air France has said that its A380s will be used on its Paris to Montreal and New York routes.

As of October 2007 Airbus had assembled 23 A380s, and was expecting the first aircraft equipped with the new electrical system (which replaces the root cause of the massive programme delays), MSN026, to be ready for 'power-on' in early 2008.

Design


A380 cabin cross section, showing economy class seating

The new Airbus is sold in two models. The **A380-800** was originally designed to carry 555 passengers in a three-class configuration or 853 passengers (538 on the main deck and 315 on the upper deck) in a single-class economy configuration. In May 2007, Airbus began marketing the same aircraft to customers with 30 fewer passengers (now 525 passengers) traded for 200 nmi more range, to better reflect trends in premium class accommodation. The design range for the -800 model is 15,200 km (8,200 nmi). The second model, the **A380-800F** freighter, will carry 150 tonnes of cargo 10,400 km (5,600 nmi). Future variants may include an **A380-900** stretch seating about 656 passengers (or up to 960 passengers in an all economy configuration) and an extended range version with the same passenger capacity as the A380-800.

The A380's wing is sized for a Maximum Take-Off Weight (MTOW) over 650 tonnes in order to accommodate these future versions, albeit with some strengthening required. The stronger wing (and structure) is used on the A380-800F freighter. This common design approach sacrifices some fuel efficiency on the A380-800 passenger model, but Airbus estimates that the size of the aircraft, coupled with the advances in technology described below, will provide lower operating costs per passenger than all current variants of Boeing 747. The A380 also features wingtip fences similar to those found on the A310 and A320 to alleviate the effects of wake turbulence, increasing fuel efficiency and performance.

Flight deck

Airbus used similar cockpit layout, procedures and handling characteristics to those of other Airbus aircraft, to reduce crew training costs. Accordingly, the A380 features an improved glass cockpit, and fly-by-wire flight controls linked to side-sticks. The improved cockpit displays feature eight 15-by-20 cm (6-by-8-inch) liquid crystal displays, all of which are physically identical and interchangeable. These comprise two Primary Flight Displays, two navigation displays, one engine parameter display, one system display and two Multi-Function Displays. These MFDs are new with the A380, and provide an easy-to-use interface to the flight management system—replacing three multifunction control and display units. They include QWERTY keyboards and trackballs, interfacing with a graphical " point-and-click" display navigation system.


The flight deck

Engines

The A380 can be fitted with two different types of engines: A380-841, A380-842 and A380-843F with Rolls-Royce Trent 900, and the A380-861 and A380-863F with Engine Alliance GP7000 turbofans. The Trent 900 is a derivative of the Trent 800, and the GP7000 has roots from the GE90 and PW4000. The Trent 900 core is a scaled version of the Trent 500, but incorporates the swept fan technology of the stillborn Trent 8104. The GP7200 has a GE90-derived core and PW4090-derived fan and low-pressure turbo-machinery. Only two of the four engines are fitted with reverse thrusters.


A Rolls-Royce Trent 900 engine on the wing of an Airbus A380

Noise reduction was an important requirement in the A380's design, and particularly affects engine design. Both engine types allow the aircraft to achieve QC/2 departure and QC/0.5 arrival noise limits under the Quota Count system set by London Heathrow Airport, which is expected to become a key destination for the A380.

Advanced materials

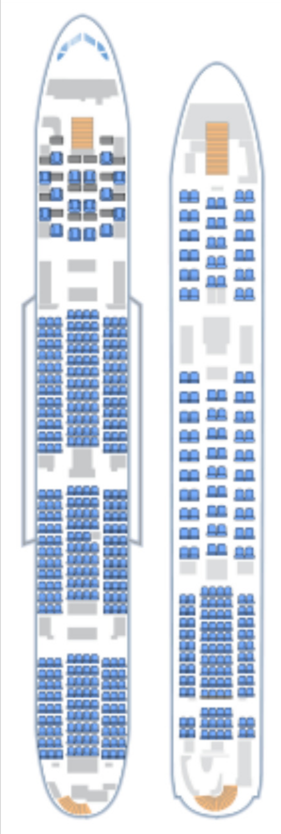
Whilst most of the fuselage is aluminium, composite materials make up 25% of the A380's airframe, by weight. Carbon-fibre reinforced plastic, glass-fibre reinforced plastic and quartz-fibre reinforced plastic are used extensively in wings, fuselage sections (such as the undercarriage and rear end of fuselage), tail surfaces, and doors. The A380 is the first commercial airliner with a central wing box made of carbon fibre reinforced plastic, and it is the first to have a wing cross-section that is smoothly contoured. Other commercial airliners have wings that are partitioned span-wise in sections. The flowing, continuous cross-section allows for maximum aerodynamic efficiency. Thermoplastics are used in the leading edges of the slats. The new material GLARE (GLAss-REinforced fibre metal laminate) is used in the upper fuselage and on the stabilizers' leading edges. This aluminium-glass-fibre laminate is lighter and has better corrosion and impact resistance than conventional aluminium alloys used in aviation. Unlike earlier composite materials, it can be repaired using conventional aluminium repair techniques. Newer weldable aluminium alloys are also used. This enables the widespread use of laser beam welding manufacturing techniques — eliminating rows of rivets and resulting in a lighter, stronger structure.

Avionics architecture

The A380 employs an Integrated Modular Avionics (IMA) architecture, first used in advanced military aircraft such as the F-22 Raptor and the Eurofighter Typhoon. It is based on a commercial off-the-shelf (COTS) design. Many previous dedicated single-purpose avionics computers are replaced by dedicated software housed in onboard processor modules and servers. This cuts the number of parts, provides increased flexibility without resorting to customised avionics, and reduces costs by using commercially available computing power. Together with IMA, the A380 avionics are very highly networked. The data communication networks use Avionics Full-Duplex Switched Ethernet, following the ARINC 664 standard. The data networks are switched, full-duplexed, star-topology and based on 100baseTX fast-Ethernet. This reduces the amount of wiring required and minimizes latency. The Network Systems Server (NSS) is the heart of A380 paperless cockpit. It eliminates the bulky manuals and charts traditionally carried by the pilots. The NSS has enough inbuilt robustness to do away with onboard backup paper documents. The A380's network and server system stores data and offers electronic documentation, providing a required equipment list, navigation charts, performance calculations, and an aircraft logbook. All are accessible to the pilot from two additional 27 cm (11 inch) diagonal LCDs, each controlled by its own keyboard and control cursor device mounted in the foldable table in front of each pilot.

Systems

Power-by-wire flight control actuators are used for the first time in civil service, backing up the primary hydraulic flight control actuators. During certain maneuvers, they augment the primary actuators. They have self-contained hydraulic and electrical power supplies. They are used as electro-hydrostatic actuators (EHA) in the aileron and elevator, and as electrical backup hydrostatic actuators (EBHA) for the rudder and some spoilers.


The A380-800 layout with 550 seats displayed

The aircraft's 350 bar (35 MPa or 5,000 psi) hydraulic system is an improvement over the typical 210 bar (21 MPa or 3,000 psi) system found in other commercial aircraft since the 1940s. First used in military aircraft, higher pressure hydraulics reduce the size of pipelines, actuators and other components for overall weight reduction. The 350 bar pressure is generated by eight de-clutchable hydraulic pumps. Pipelines are typically made from titanium and the system features both fuel and air-cooled heat exchangers. The hydraulics system architecture also differs significantly from other airliners. Self-contained electrically powered hydraulic power packs, instead of a secondary hydraulic system, are the backups for the primary systems. This saves weight and reduces maintenance.

The A380 uses four 150 kVA variable-frequency electrical generators eliminating the constant speed drives for better reliability. The A380 uses aluminium power cables instead of copper for greater weight savings due to the number of cables used for an aircraft of this size and complexity. The electrical power system is fully computerized and many contactors and breakers have been replaced by solid-state devices for better performance and increased reliability.

The A380 features a bulbless illumination system. LEDs are employed in the cabin, cockpit, cargo and other fuselage areas. The cabin lighting features programmable multi-spectral LEDs capable of creating a cabin ambience simulating daylight, night or shades in between. On the outside of the aircraft, HID lighting is used to give brighter, whiter and better quality illumination. These two technologies provide brightness and a service life superior to traditional incandescent light bulbs.

The A380 was initially planned without thrust reversers, as Airbus believed it to have ample braking capacity. The FAA disagreed, and Airbus elected to fit only the two inboard engines with them. The two outboard engines do not have reversers, reducing the amount of debris stirred up during landing. The A380 features electrically actuated thrust reversers, giving them better reliability than their pneumatic or hydraulic equivalents, in addition to saving weight.

Passenger provisions

The A380 produces 50% less cabin noise than a 747 and has higher cabin air pressure (equivalent to an altitude of 1500 metres (5000 feet) versus 2500 metres (8000 feet)); both features are expected to reduce the effects of travel fatigue. The upper and lower decks are connected by two stairways, fore and aft, wide enough to accommodate two passengers side-by-side. In a 555-passenger configuration, the A380 has 33% more seats than a 747-400 in a standard three-class configuration but 50% more cabin area and volume, resulting in more space per passenger. Its maximum certified carrying capacity is 853 passengers in an all-economy-class configuration.

Economy class on the first Singapore Airlines aircraft

Compared to a 747, the A380 has larger windows and overhead bins, and 60 cm (2 feet) of extra headroom. The wider cabin allows for 48 cm (19 inch) wide economy seats instead of 43 cm (17 inch) seats on a 747, although the seat pitch of 81 cm (32 inch) is the same as that on a 747. Singapore Airline's economy-class seats feature 27 cm (10.6 inch) LCD screens in each seatback, as well as an AC power supply in most seats; business-class seats are 84 cm (34 inches) wide, can lay flat for sleeping, and have 39 cm (15.4 inch) LCD screens.

Airbus' initial publicity stressed the comfort and space of the A380's cabin, anticipating installations such as relaxation areas, bars, duty-free shops, and beauty salons. Virgin Atlantic Airways already offers a bar as part of its "Upper Class" service on its A340 and 747 aircraft, and has announced plans to include casinos, double beds, and gymnasiums on its A380s. Singapore Airlines offers twelve fully-enclosed first-class suites on its A380, each with one full and one secondary seat, full-sized bed, desk, personal storage, and 58-cm (23-inch) LCD screen at a 20% to 25% price premium over standard first class seating. Four of these suites are in the form of two "double" suites featuring a double bed. Emirates has not yet revealed their front-end A380 product although Qantas Airways has shown their product which features a long flat-bed that converts from the seat but does not have privacy doors.

Production

Major structural sections of the A380 are built in France, Germany, Spain, and the United Kingdom. Due to their size, they are brought to the assembly hall in Toulouse in France by surface transportation, rather than by the A300-600ST *Beluga* aircraft used for other Airbus models. Components of the A380 are provided by suppliers from around the world; the five largest contributors, by value, are Rolls-Royce, SAFRAN, United Technologies, General Electric, and Goodrich.

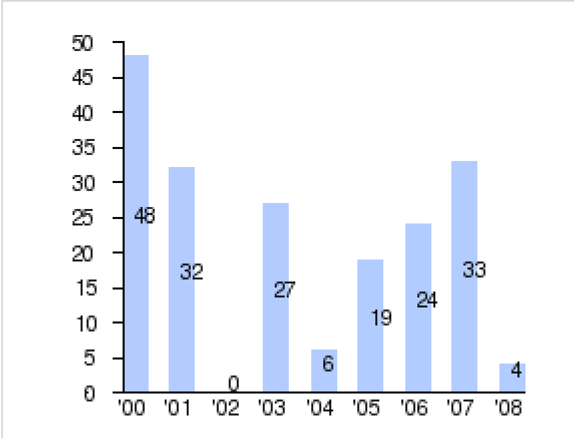
A380 transporter ship *Ville de Bordeaux*

The front and rear sections of the fuselage are loaded on an Airbus Roll-on/roll-off (RORO) ship, *Ville de Bordeaux*, in Hamburg in northern Germany, whence they are shipped to the United Kingdom. The wings, which are manufactured at Filton in Bristol and Broughton in North Wales, are transported by barge to Mostyn docks, where the ship adds them to its cargo. In Saint-Nazaire in western France, the ship trades the fuselage sections from Hamburg for larger, assembled sections, some of which include the nose. The ship unloads in Bordeaux. Afterwards, the ship picks up the belly and tail sections by Construcciones Aeronáuticas SA in Cádiz in southern Spain, and delivers them to Bordeaux. From there, the A380 parts are transported by barge to Langon, and by oversize road convoys to the assembly hall in Toulouse. New wider roads, canal systems and barges were developed to deliver the A380 parts. After assembly, the aircraft are flown to Hamburg, XFW to be furnished and painted. It takes 3,600 litres (950 gallons) of paint to cover the 3,100 m² (33,000 ft²) exterior of an A380.

Airbus sized the production facilities and supply chain for a production rate of four A380s per month.

Orders and deliveries

A380-800 orders, by year



Eighteen customers have ordered the A380, including an order from aircraft lessor ILFC. Total orders for the A380 stand at 193, of which 189 were firm as of 31 December 2007. Orders for the freighter model reached 27 but dwindled to zero following the production delays. Airbus expects to sell a total of 750 aircraft, and estimated break-even at 420 units, increased from 270 due to the delays and the falling exchange rate of the US dollar. In April 2007, Airbus CEO Louis Gallois said that break-even had risen further, but declined to give the new figure. Industry analysts anticipate between 400 and 880 sales by 2025. As of 2006, the list price of an A380 was US$ 296 to 316 million, depending on equipment installed.

The first private buyer of an A380 for personal use is Saudi Prince Alwaleed bin Talal, who reportedly spent only fifteen minutes on the plane before deciding to put one under contract.

Industry sources have stated that the United States Air Force Air Mobility Command is looking into possibly purchasing the A380 as a replacement for the aging Boeing 747s in the role of presidential transport. The USAF may also be interested in a military version of the A380F as a tactical transport aircraft, replacing the C-5 Galaxy.

Deliveries

2007	2008	2009	2010	Total
1	**1** *(13)*	*(25)*	*(44)*	**2** *(83)*


Emirates Airline has placed the most orders so far. (A380 F-WWDD in the airline's livery at the 2005 Dubai Airshow)

Anticipated yearly totals are in parentheses.

Technical concerns

Several concerns about the A380 have arisen during its development. Airbus has addressed these concerns as required to obtain a type certificate from the European Aviation Safety Agency and its American counterpart, the Federal Aviation Administration.

Ground operations

Early critics claimed that the A380 would damage taxiways and other airport surfaces. However, the pressure exerted by its wheels is lower than that of a Boeing 747 or Boeing 777 because the A380 has 22 wheels, four more than the 747, and eight more than the 777. Airbus measured pavement loads using a 540-tonne (595 short tons) ballasted test rig, designed to replicate the landing gear of the A380. The rig was towed over a section of pavement at Airbus' facilities that had been instrumented with embedded load sensors.


The A380's 20-wheel main landing gear

Based on its wingspan, the U.S. FAA classifies the A380 as a Design Group VI aircraft, and originally required a width of 60 m (200 ft) for runways and 30 m (100 ft) for taxiways, compared with 45 m (150 ft) and 23 m (75 ft) for Design Group V aircraft such as the Boeing 747. The FAA also considered limiting the taxi speed of the A380 to 25 km/h (15 mph) when operating on Group V infrastructure, but issued waivers related to the speed restriction and some of the proposed runway widening requirements. Airbus claimed from the beginning that the A380 could safely operate on Group V runways and taxiways, without the need for widening. In July 2007, the FAA and EASA agreed to let the A380 operate on 45 m runways without restrictions. The International Civil Aviation Organization (ICAO) is still disputing this issue.

The A380 was designed to fit within an 80 × 80 m airport gate, and can land or take off on any runway that can accommodate a Boeing 747. Its large wingspan can require some taxiway and apron reconfigurations, to maintain safe separation margins when two of the aircraft pass each other. Taxiway shoulders may be required to be paved to reduce the likelihood of foreign object damage caused to (or by) the outboard engines, which overhang more than 25 m (80 ft) from the centre line of the aircraft. Any taxiway or runway bridge must be capable of supporting the A380's maximum weight. The terminal gate must be sized such that the A380's wings do not block adjacent gates, and may also provide multiple jetway bridges for simultaneous boarding on both decks.

Service vehicles with lifts capable of reaching the upper deck should be obtained, as well as tractors capable of handling the A380's maximum ramp weight. The A380 test aircraft have participated in a campaign of airport compatibility testing to verify the modifications already made at several large airports, visiting a number of airports around the world.

A380 being serviced by three separate jetways at Frankfurt Airport; two for the main deck and one for the upper deck.

Wake turbulence

The A380 generates more wake turbulence on takeoff and landing than existing aircraft types, requiring increased airport approach and departure spacing for following aircraft.

In 2005, the ICAO recommended that provisional separation criteria for the A380 be substantially greater than for the 747 because preliminary flight test data suggested a stronger wake than the 747. These criteria were in effect while the A380 Wake Vortex Steering Group, with representatives from the JAA, Eurocontrol, the FAA, and Airbus, refined its 3-year study of the issue with additional flight testing. In September 2006, the working group presented its conclusions to the ICAO, which rendered final guidance on the issue in November 2006. The working group concluded that an aircraft trailing an A380 during approach needs to maintain a separation of 6 nmi, 8 nmi and 10 nmi respectively for ICAO "Heavy", "Medium", and "Light" aircraft categories, compared with 4 nmi, 5 nmi and 6 nmi spacing for other heavy aircraft. However, the working group found no need to limit the A380's trailing distance behind another aircraft, potentially making up for some of the increased spacing behind the A380. On departure behind an A380, the working group concluded that "Heavy" aircraft are required to wait two minutes, and "Medium"/"Light" aircraft three minutes for time based operations. Finally, the working group did not recommend any modified restrictions on vertical or horizontal separation criteria during cruise.

During the A380's maiden trip to the United States in 2007, air traffic control used the callsign suffix "Super" to distinguish the A380 from " Heavy" aircraft.

Wing strength

During the destructive wing strength certification test on MSN5000, the test wing of the A380 failed to meet the certification requirement of 150% of limit load. Limit load is the maximum load expected during operation in the design life of an aircraft. The test wing buckled between the inboard and outboard engines at 147% of limit load, as the wing tip reached a vertical deflection of 7.4 m (24.3 ft). Airbus initially stated that the test article represented an early design, and that the load requirement would be verified by analysis of changes already made. Subsequently, Airbus announced that modifications adding 30 kg to the wing would be made to provide the required strength.

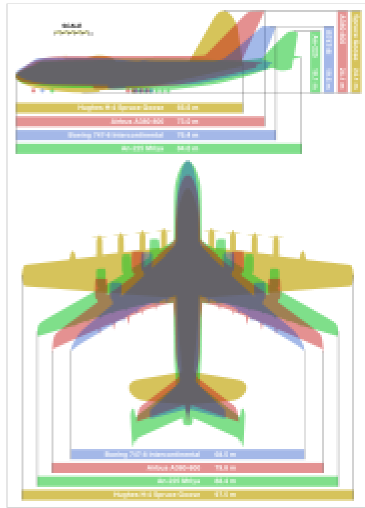
Future versions

Airbus A380-900

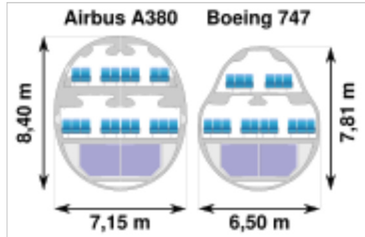
Airbus top sales executive and COO John Leahy confirmed the plans for an enlarged version, the A380-900. This version would have a seat capacity of 650 passengers in standard configuration, and of around 900 passengers in economy-only configuration. The development of the A380-900 is planned to start once the regular production of the A380-800 variant reaches 40 planes per year. Airbus foresees to reach this production capacity in 2010. Given this timeline, the first A380-900s could be delivered to customers around 2015, hence at about the same time as the freighter variant A380-800F. Airlines, including Emirates , Virgin Atlantic , and Cathay Pacific , along with leasing company ILFC have already expressed great interest in the extended model. According to an interview in Airliner World magazine's December issue, Singapore Airlines CEO Chew Choon Seng revealed at the delivery of their first A380-800 that the airline is keeping their options open with their order, by only defining their first ten A380s as -800s. The remaining 9 aircraft could be in fact be transferred to -900s.

Specifications

Measurement	A380-800	A380-800F
Cockpit crew	Two	
Seating capacity	525 (3-class) 644 (2-class) 853 (1-class)	12 couriers
Length	73 m (239 ft 6 in)	
Span	79.8 m (261 ft 10 in)	
Height	24.1 m (79 ft 1 in)	
Wheelbase	30.4 m (99 ft 8 in)	
Outside fuselage width	7.14 m (23 ft 6 in)	
Cabin width, main deck	6.60 m (21 ft 8 in)	
Cabin width, upper deck	5.94 m (19 ft 6 in)	
Wing area	845 m² (9,100 sq ft)	
Operating empty weight	276,800 kg (610,200 lb)	252,200 kg (556,000 lb)
Maximum take-off weight	560,000 kg (1,235,000 lb)	590,000 kg (1,300,000 lb)
Maximum payload	90,800 kg (200,000 lb)	152,400 kg (336,000 lb)
Cruising speed	Mach 0.85	
Maximum cruising speed	Mach 0.89	
Maximum speed	Mach 0.96	
Take off run at MTOW	2,750 m (9,020 ft)	2,900 m (9,510 ft)
Range at design load	15,200 km (8,200 nmi)	10,400 km (5,600 nmi)
Service ceiling	13,115 m (43,000 ft)	
Maximum fuel capacity	310,000 L (81,890 US gal)	310,000 L (81,890 US gal), 356,000 L (94,000 US gal) option
Engines (4 x)	GP7270 (A380-861) Trent 970/B (A380-841) Trent 972/B (A380-842)	GP7277 (A380-863F) Trent 977/B (A380-843F)



Size comparison between four of the largest aircraft. Airbus A380 (red), Boeing 747-8I (blue), Antonov An-225 (green) and Hughes H-4 (yellow). Click to enlarge.



Economy class fuselage-comparison between Airbus A380 and the front-section of Boeing 747, the next-largest passenger aircraft

Aircraft carrier

2008/9 Schools Wikipedia Selection. Related subjects: Air & Sea transport

An **aircraft carrier** is a warship designed to deploy and recover aircraft, acting as a sea-going airbase. Aircraft carriers thus allow a naval force to project air power great distances without having to depend on local bases for staging aircraft operations. They have evolved from wooden vessels used to deploy a balloon into nuclear powered warships that carry dozens of fixed and rotary wing aircraft.

Four aircraft carriers, (bottom-to-top) *Principe de Asturias*, amphibious assault ship USS *Wasp*, USS *Forrestal* and light V/STOL carrier HMS *Invincible*, showing size differences of late 20th century carriers

Balloon carriers were the first ships to deploy manned aircraft, used during the 19th and early 20th century, mainly for observation purposes. The 1903 advent of fixed wing airplanes was followed in 1910 by the first flight of such an aircraft from the deck of a US Navy cruiser. Seaplanes and seaplane tender support ships, such as HMS *Engadine*, followed. The development of flat top vessels produced the first large fleet ships. This evolution was well underway by the mid 1920s, resulting in ships such as the *HMS Hermes*, *Hōshō*, and the Lexington class aircraft carriers.

World War II saw the first large scale use and further refinement of the aircraft carrier, spawning several types. Escort aircraft carriers, such as USS *Barnes*, were built only during World War II. Although some were purpose built, most were converted from merchant ships, and were a stop-gap measure in order to provide air support for convoys and amphibious invasions. Light aircraft carriers, such as USS *Independence* represented a larger, more "militarized" version of the escort carrier concept. Although the light carriers usually carried the same size air groups as escort carriers, they had the advantage of higher speed as they had been converted from cruisers under construction rather than civilian merchant ships.

Wartime emergencies also saw the creation or conversion of other, unconventional aircraft carriers. CAM ships, like the SS *Michael E*, were cargo carrying merchant ships which could launch but not retrieve fighter aircraft from a catapult. These vessels were an emergency measure during World War II as were Merchant aircraft carriers (MACs), such as MV *Empire MacAlpine*, another emergency measure which saw cargo-carrying merchant ships equipped with flight decks. Battlecarriers were created by the Imperial Japanese Navy to partially compensate for the loss of carrier strength at Midway. Two of them were made from *Ise* class battleships during late 1943. The aft turrets were removed and replaced with a hangar, deck and catapult. The heavy cruiser *Mogami* concurrently received a similar conversion. This "half and half" design was an unsuccessful compromise, being neither one thing nor the other. Submarine aircraft carriers, such as the French *Surcouf*, or the Japanese I-400 class submarines, which were capable of carrying 3 Aichi M6A *Seiran* aircraft, were first built in the 1920s, but were generally unsuccessful at war. Modern navies that operate such ships treat aircraft carriers as the capital ship of the fleet, a role previously played by the battleship. The change, part of the growth of air power as a significant part of warfare, took place during World War II. This change was driven by the superior range, flexibility and effectiveness of carrier-launched aircraft.

Following the war, the scope of carrier operations continued to increase in size and importance. The Supercarrier, typically displacing 75,000 tonnes or greater has been the pinnacle of carrier development since their introduction. Most are powered by nuclear reactors and form the core of a fleet designed to operate far from home. Amphibious assault carriers, such as USS *Tarawa* or HMS *Ocean*, which serve the purpose of carrying and landing Marines and operate a large contingent of helicopters for that purpose. They have a secondary capability to operate VSTOL aircraft. Also known as "commando carriers" or "helicopter carriers".

Lacking the firepower of other warships, carriers by themselves are considered vulnerable to attack by other ships, aircraft, submarines or missiles and therefore travel as part of a carrier battle group (CVBG) for their protection. Unlike other types of capital ships in the 20th century, aircraft carrier designs since World War II have been effectively unlimited by any consideration save budgetary, and the ships have increased in size to handle the larger aircraft: The large, modern *Nimitz* class of United States Navy carriers has a displacement nearly four times that of the World War II-era USS *Enterprise* yet its complement of aircraft is roughly the same, a consequence of the steadily increasing size of military aircraft over the years.

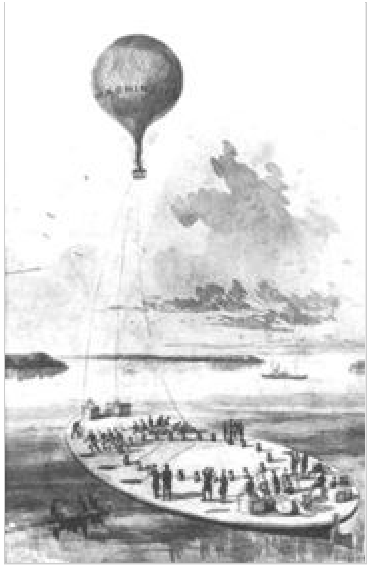
History and milestones

Though aircraft carriers are given their definition with respect to fixed-wing aircraft, the first known instance of using a ship for airborne operations occurred in 1806, when the British Royal Navy's Lord Thomas Cochrane launched kites from the 32-gun frigate *HMS Pallas* in order to drop propaganda leaflets on the French territory.

Balloon carriers

On July 12, 1849, the Austrian Navy ship *Vulcano* launched a manned hot air balloon in order to drop bombs on Venice, although the attempt failed due to contrary winds.

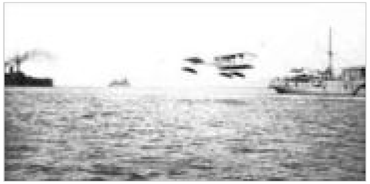
Later, during the American Civil War, about the time of the Peninsula Campaign, gas-filled balloons were being used to perform reconnaissance on Confederate positions. The battles soon turned inland into the heavily forested areas of the Peninsula, however, where balloons could not travel. A coal barge, the George Washington Parke Custis, was cleared of all deck rigging to accommodate the gas generators and apparatus of balloons. From the GWP Prof. Thaddeus S. C. Lowe, Chief Aeronaut of the Union Army Balloon Corps, made his first ascents over the Potomac River and telegraphed claims of the success of the first aerial venture ever made from a water-borne vessel. Other barges were converted to assist with the other military balloons transported about the eastern waterways. It is only fair to point out in deference to modern aircraft carriers that none of these Civil War crafts had ever taken to the high seas.



The Union Army balloon *Washington* aboard the *George Washington Parke Custis*, towed by the tug *Coeur de Lion*.

Balloons launched from ships led to the development of balloon carriers, or balloon tenders, during World War I, by the navies of Great Britain, France, Germany, Italy, Russia, and Sweden. About ten such "balloon tenders" were built, their main objective being aerial observation posts. These ships were either decommissioned or converted to seaplane tenders after the war.

Seaplane carriers



The first seaplane carrier, the French *La Foudre* (right, with hangar and crane), with one of her *Canard Voisin* seaplanes taking off, during tactical exercises in June 1912.

The invention of the seaplane in March 1910 with the French *Le Canard* led to the earliest development of a ship designed to carry airplanes, albeit equipped with floats: in December 1911 appears the French Navy *La Foudre*, the first seaplane carrier, and the first known carrier of airplanes. Commissioned as a seaplane tender, and carrying float-equipped planes under hangars on the main deck, from where they were lowered on the sea with a crane, she participated in tactical exercises in the Mediterranean in 1912. *La Foudre* was further modified in November 1913 with a 10 meter long flat deck to launch her seaplanes.

HMS *Hermes*, temporarily converted as an experimental seaplane carrier in April-May 1913, is also one of the first seaplane carriers, and the first experimental seaplane carrier of the British Navy. She was originally laid down as a merchant ship, but was converted on the building stocks to be a seaplane carrier for a few trials in 1913, before being converted again to a cruiser, and back again to a seaplane carrier in 1914. She was sunk by a German submarine in October 1914. The first seaplane tender of the US Navy was the USS *Mississippi*, converted to that role in December 1913.

Many cruisers and capital ships of the inter-war years often carried a catapult launched seaplane for reconnaissance and spotting the fall of the guns. It was launched by a catapult and recovered by crane from the water after landing. These were highly successful during World War II; there were many notable successes early in the war as shown by HMS *Warspite*'s float equipped Swordfish during operations in the Norwegian fjords in 1940. The Japanese Rufe floatplane derived from the Zero was a formidable fighter with only a slight loss in flight performance, one of their pilots scored 26 kills in the A6M2-N Rufe; a score only bettered by a handful of American pilots throughout WW2. Other Japanese seaplanes launched from tenders and warships sank merchant ships and small-scale ground attacks. The culmination of the type was the American 300+ mph (480 km/h) Curtiss SC Seahawk which was actually a fighter aircraft like the Rufe in addition to a two-seat gunnery spotter and transport for an injured man in a litter. Spotter seaplane aircraft on U.S. Navy cruisers and battleships were in service until 1949. Seaplane fighters were considered poor combat aircraft compared to their carrier-launched brethren; they were slower due to the drag of their pontoons or boat hulls. Contemporary propeller-driven, land-based fighter aircraft were much faster (450-480 mph / 720-770 km/h as opposed to 300-350 mph / 480-560 km/h) and more heavily armed. The Curtiss Seahawk only had two 0.50 inch (12.7 mm) calibre machine guns compared to four 20 mm cannon in the Grumman F8F Bearcat or four 0.50 (12.7 mm) cal machine guns plus two 20 mm cannon in the Vought F4U Corsair. Jet aircraft of just a few years later were faster still (500+ mph) and still better armed, especially with the development of air to air missiles in the early to mid 1950s.

Genesis of the flat-deck carrier

As heavier-than-air aircraft developed in the early 20th century various navies began to take an interest in their potential use as scouts for their big gun warships. In 1909 the French inventor Clément Ader published in his book " L'Aviation Militaire" the description of a ship to operate airplanes at sea, with a flat flight deck, an island superstructure, deck elevators and a hangar bay. That year the US Naval Attaché in Paris sent a report on his observations.

> "An airplane-carrying vessel is indispensable. These vessels will be constructed on a plan very different from what is currently used. First of all the deck will be cleared of all obstacles. It will be flat, as wide as possible without jeopardizing the nautical lines of the hull, and it will look like a landing field."
>
> Clément Ader, " L'Aviation Militaire", 1909


Ely takes off from USS *Birmingham*, 14 November 1910.

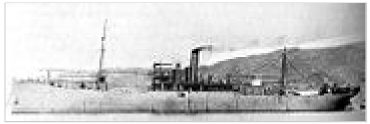
A number of experimental flights were made to test the concept. Eugene Ely was the first pilot to launch from a stationary ship in November 1910. He took off from a structure fixed over the forecastle of the US armored cruiser USS *Birmingham* at Hampton Roads, Virginia and landed nearby on Willoughby Spit after some five minutes in the air.

On January 18, 1911 he became the first pilot to land on a stationary ship. He took off from the Tanforan racetrack and landed on a similar temporary structure on the aft of USS *Pennsylvania* anchored at the San Francisco waterfront — the improvised braking system of sandbags and ropes led directly to the arrestor hook and wires described above. His aircraft was then turned around and he was able to take off again. Commander Charles Samson, RN, became the first airman to take off from a moving warship on May 2, 1912. He took off in a Short S27 from the battleship HMS *Hibernia* while she steamed at 10.5 knots (19 km/h) during the Royal Fleet Review at Weymouth.

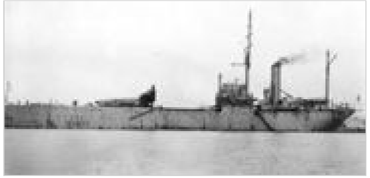

Ely lands on USS *Pennsylvania*, 18 January 1911.

World War I

The first strike from a carrier against a land target as well as a sea target took place in September 1914 when the Imperial Japanese Navy seaplane carrier *Wakamiya* conducted the world's first naval-launched air raids from Kiaochow Bay during the Battle of Tsingtao in China. The four Maurice Farman seaplanes bombarded German-held land targets (communication centers and command centers) and damaged a German minelayer in the Tsingtao peninsula from September until November 6, 1914, when the Germans surrendered. On the Western front the first naval air raid occurred on December 25, 1914 when twelve seaplanes from HMS *Engadine*, *Riviera* and *Empress* (cross-channel steamers converted into seaplane carriers) attacked the Zeppelin base at Cuxhaven. The attack was not a complete success, although a German warship was damaged; nevertheless the raid demonstrated in the European theatre the feasibility of attack by ship-borne aircraft and showed the strategic importance of this new weapon.



The Japanese seaplane carrier *Wakamiya* conducted the world's first naval-launched air raids in September 1914.



HMS *Ark Royal*, a seaplane carrier also equipped with two regular aeroplanes, was arguably the first modern aircraft carrier.

HMS *Ark Royal* was arguably the first modern aircraft carrier. She was originally laid down as a merchant ship, but was converted on the building stocks to be a hybrid airplane/seaplane carrier with a launch platform. Launched September 5, 1914, she served in the Dardanelles campaign and throughout World War I.

Other carrier operations were mounted during the war the most successful taking place on 19 July 1918 when seven Sopwith Camels launched from HMS *Furious* attacked the German Zeppelin base at Tondern, with two 50 lb (23 kg) bombs each. Several airships and balloons were destroyed, but as the carrier had no method of recovering the aircraft safely, two of the pilots ditched their aircraft in the sea alongside the carrier while the others headed for neutral Denmark.

Inter-war years

The Washington Naval Treaty of 1922 placed strict limits on the tonnages of battleships and battlecruisers for the major naval powers after World War I, as well as limits not only on the total tonnage for carriers, but also an upper limit on 27,000 tonnes for each ship. Although exceptions were made regarding the max ship tonnage (fleet units counted, experimental units did not), the total tonnage could not be exceeded. However, while all of the major navies were over-tonnage on battleships, they were all considerably under-tonnage on aircraft carriers. Consequently, many battleships and battlecruisers under construction (or in service) were converted into aircraft carriers. The first ship to have a full length flat deck was HMS *Argus* the conversion of which was completed in September 1918, with the U.S. Navy not following suit until 1920, when the conversion of USS *Langley* (an experimental ship which did not count against America's carrier tonnage) had completed. The first American fleet carriers would not join the service until November, 1927 when the USS *Saratoga* was commissioned. (USS *Lexington* was commissioned in December of that year.)

The first full-length flat deck, HMS *Argus* in 1918


The Imperial Japanese Navy's 1922 *Hōshō*, was the world's first built-from-the-keel-up aircraft carrier.

The first purpose-designed aircraft carrier to be laid down was the HMS *Hermes* in 1918, the next year Japan began work on *Hōshō*. Three years later in December 1922, *Hōshō* became the first to be commissioned while HMS *Hermes* began service in July 1923. *Hermes*' design preceded and influenced that of *Hōshō*, and its construction actually began earlier, but numerous tests, experiments and budget considerations delayed its commission.

By the late 1930s, aircraft carriers around the world typically carried three types of aircraft: torpedo bombers, also used for conventional bombings and reconnaissance; dive bombers, also used for reconnaissance (in the U.S. Navy, this type of aircraft were known as "scout bombers"); and fighters for fleet defense and bomber escort duties. Because of the restricted space on aircraft carriers, all these aircraft were of small, single-engined types, usually with folding wings to facilitate storage.

World War II

Aircraft carriers played a significant role in World War II. With seven aircraft carriers afloat, the British Royal Navy had a considerable numerical advantage at the start of the war as neither the Germans nor the Italians had carriers of their own. However, the vulnerability of carriers compared to traditional battleships when forced into a gun-range encounter was quickly illustrated by the sinking of HMS *Glorious* by German battlecruisers during the Norwegian campaign in 1940.

This apparent weakness to battleships was turned on its head in November 1940 when HMS *Illustrious* launched a long-range strike on the Italian fleet at Taranto. This operation incapacitated three of the six battleships in the harbour at a cost of two of the 21 attacking Fairey Swordfish torpedo bombers. Carriers also played a major part in reinforcing Malta, both by transporting planes and by defending convoys sent to supply the besieged island. The use of carriers prevented the Italian Navy and land-based German aircraft from dominating the Mediterranean theatre.

In the Atlantic, aircraft from HMS *Ark Royal* and HMS *Victorious* were responsible for slowing *Bismarck* during May 1941. Later in the war, escort carriers proved their worth guarding convoys crossing the Atlantic and Arctic oceans.

Many of the major battles in the Pacific involved aircraft carriers. Japan started the war with ten aircraft carriers, the largest and most modern carrier fleet in the world at that time. There were six American aircraft carriers at the beginning of the hostilities, although only three of them were operating in the Pacific.

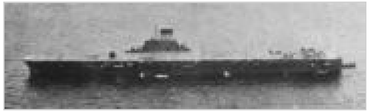
Drawing on the 1939 Japanese development of shallow water modifications for aerial torpedoes and the 1940 British aerial attack on the Italian fleet at Taranto, the 1941 Japanese surprise attack on Pearl Harbour was a clear illustration of the power projection capability afforded by a large force of modern carriers. Concentrating six flattops in a single striking unit marked a turning point in naval history, as no other nation had fielded anything comparable. (Though Germany and Italy began construction of carriers, neither were completed. Of the two, Germany's *Graf Zeppelin* had the greater potential.)

Meanwhile, the Japanese began their advance through Southeast Asia and the sinking of *Prince of Wales* and *Repulse* by Japanese land-based aircraft drove home the need for this ship class for fleet defence from aerial attack. In April 1942, the Japanese fast carrier strike force ranged into the Indian Ocean and sank shipping, including the damaged and undefended carrier HMS *Hermes*. Smaller Allied fleets with inadequate aerial protection were forced to retreat or be destroyed. In the Coral Sea, US and Japanese fleets traded aircraft strikes in the first battle where neither side's ships sighted the other. At the Battle of Midway all four Japanese carriers engaged were sunk by planes from three American carriers (one of which was lost) and the battle is considered the turning point of the war in the Pacific. Notably, the battle was orchestrated by the Japanese to draw out American carriers that had proven very elusive and troublesome to the Japanese.

Subsequently the US was able to build up large numbers of aircraft aboard a mixture of fleet, light and (newly commissioned) escort carriers, primarily with the introduction of the Essex class in 1943. These ships, around which were built the fast carrier task forces of the Third and Fifth Fleets, played a major part in winning the Pacific war. The eclipse of the battleship as the primary component of a fleet was clearly illustrated by the sinking of the largest battleship ever built, *Yamato*, by carrier-borne aircraft in 1945. Japan also built the largest aircraft carrier of the war, *Shinano*, which was a *Yamato* class ship converted mid-way through construction after the disastrous loss of four fleet carriers at Midway. She was sunk by a patrolling US submarine while in transit shortly after commissioning, but before being fully outfitted or operational in November 1944.

4 US Navy carriers right after the war, showing the size and length difference between an early battlecruiser conversion, the *Saratoga* (bottom), an early fleet carrier Enterprise (2nd from bottom), a war time built *Essex*-class carrier (2nd from top, the *Hornet*) and a light carrier based on a cruiser hull, the *San Jacinto* (top).

Important innovations just before and during World War II



Japanese carrier *Taihō* had a hurricane bow.

USS Saratoga circa 1935.

Hurricane bow

A hurricane bow is a completely enclosed hangar deck, first seen on the American Lexington class aircraft carriers which entered service in 1927. Combat experience proved it to be by far the most useful configuration for the bow of the ship among others that were tried; including second flying-off decks and an anti-aircraft battery (the latter was the most common American configuration during World War II). This feature would be re-incorporated into American carriers post-war. The Japanese carrier *Taihō* was the first of their ships to incorporate it.

Light aircraft carriers

The loss of three major carriers in quick succession in the Pacific led the US Navy to develop the light carrier (CVL) from light cruiser hulls that had already been laid down. They were intended to provide additional fast carriers, as escort carriers did not have the requisite speed to keep up with the fleet carriers and their escorts. The actual U.S. Navy classification was small aircraft carrier (CVL), not light. Prior to July 1943, they were just classified as aircraft carriers (CV).

The British Royal Navy made a similar design which served both them and Commonwealth countries after World War II. One of these carriers, India's INS *Viraat*, formerly HMS *Hermes*, is still being used.

Escort carriers and merchant aircraft carriers

To protect Atlantic convoys, the British developed what they called Merchant Aircraft Carriers, which were merchant ships equipped with a flat deck for half a dozen aircraft. These operated with civilian crews, under merchant colors, and carried their normal cargo besides providing air support for the convoy. As there was no lift or hangar, aircraft maintenance was limited and the aircraft spent the entire trip sitting on the deck.

These served as stop-gap until dedicated escort carriers could be built in the US (US classification *CVE*). About a third of the size of a fleet carrier, it carried about two dozen aircraft for anti-submarine duties. Over one hundred were built or converted from merchantmen.

Escort carriers were built in the US from two basic hull designs: one from a merchant ship, and the other from a slightly larger, slightly faster tanker. Besides defending convoys, these were used to transport aircraft across the ocean. Nevertheless, some participated in the battles to liberate the Philippines, notably the Battle off Samar in which six escort carriers and their escorting destroyers briefly took on five Japanese battleships and bluffed them into retreating.

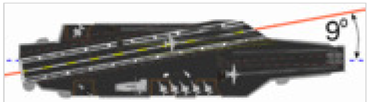
Catapult aircraft merchantmen

As an emergency stop-gap before sufficient merchant aircraft carriers became available, the British provided air cover for convoys using *Catapult aircraft merchantman* (CAM ships) and merchant aircraft carriers. CAM ships were merchant vessels equipped with an aircraft, usually a battle-weary Hawker Hurricane, launched by a catapult. Once launched, the aircraft could not land back on the deck and had to ditch in the sea if it was not within range of land. Over two years, fewer than 10 launches were ever made, yet these flights did have some success: 6 bombers for the loss of a single pilot.

Post-war developments

Three major post-war developments came from the need to improve operations of jet-powered aircraft, which had higher weights and landing speeds than their propeller-powered forbears. The first jets were tested as early as 3 December 1945; a de Havilland Vampire and jets were operating by the early 1950s from carriers.

Angled decks



The angled flight deck allows for safe simultaneous launch and recovery of aircraft.

During the Second World War, aircraft would land on the flight deck parallel to the long axis of the ship's hull. Aircraft which had already landed would be parked on the deck at the bow end of the flight deck. A crash barrier was raised behind them to stop any landing aircraft which overshot the landing area because its landing hook missed the arrestor cables. If this happened, it would often cause serious damage or injury and even, if the crash barrier was not strong enough, destruction of parked aircraft.

An important development of the early 1950s was the British invention of the angled deck, where the runway was canted at an angle of a few degrees across the ship. If an aircraft misses the arrestor cables, the pilot only needs to increase engine power to maximum to get airborne again (referred to as " boltering") and will not hit the parked aircraft because the angled deck points out over the sea.

Steam catapults

The modern steam-powered catapult, powered by steam from the ship's boilers or reactors, was invented by Commander C.C. Mitchell of the British RNVR. It was widely adopted following trials on HMS *Perseus* between 1950 and 1952 which showed it to be more powerful and reliable than the compressed air catapults which had been introduced in the 1940s.

Landing system

Another British invention was the Mirror Landing Aid. This was a gyroscopically-controlled convex mirror (in later designs replaced by a Fresnel lens) on the port side of the deck. Either side of the mirror was a line of green coloured lights, the "datum lights". A bright orange light was shone into the mirror creating the "ball" (or "meatball" in later USN parlance) which could be seen by the aviator who was about to land. The position of the ball compared to the datum lights indicated the aircraft's position in relation to the desired glidepath: if the ball was above the datum, the plane was high; below the datum, the plane was low; between the datum, the plane was on glidepath. The gyro stabilisation compensated for the movement of the flight deck due to the sea, giving a constant glidepath.

Landing optics of *Charles de Gaulle*, note that this system is of the later Fresnel lens design.

Nuclear age

The US Navy attempted to become a strategic nuclear force in parallel with the USAF long range bombers with the project to build *United States*, which was termed CVA, with the "A" signifying "atomic". This ship would have carried long range twin-engine bombers, each of which could carry an atomic bomb. The project was canceled under pressure from the newly-created United States Air Force, and the letter "A" was re-cycled to mean "attack." But this only delayed the growth of carriers. Nuclear weapons would be part of the carrier weapons load despite Air Force objections beginning in 1955 aboard USS *Forrestal*, and by the end of the fifties the Navy had a series of nuclear-armed attack aircraft (see also USS *Franklin D. Roosevelt* (CV-42)).

The US Navy also built the first aircraft carrier to be powered by nuclear reactors. USS *Enterprise* is powered by eight nuclear reactors and was the second surface warship (after USS *Long Beach*) to be powered in this way. Subsequent supercarriers starting with USS *Nimitz* took advantage of this technology to increase their endurance utilizing only two reactors. The only other nation to have followed the US lead is France with *Charles de Gaulle* although nuclear power is used for submarine propulsion by France, Great Britain, China and the former Soviet Union.

Helicopters

The post-war years also saw the development of the helicopter, with a variety of useful roles and mission capability aboard aircraft carriers. Whereas fixed-wing aircraft are suited to air-to-air combat and air-to-surface attack, helicopters are used to transport equipment and personnel and can be used in an anti-submarine warfare (ASW) role, with dipping sonar, air-launched torpedoes, and depth charges; as well as anti-surface vessel warfare, with air-launched anti-ship missiles.


The *Tripoli*, a US Navy *Iwo Jima* class helicopter carrier

In the late 1950s and early 1960s, the UK and the U.S. converted some of their older carriers into Commando Carriers; sea-going helicopter airfields like HMS *Bulwark*. To mitigate against the expensive connotations of the term "aircraft carrier", the new *Invincible* class carriers were originally designated as "through deck cruisers" and were initially helicopter-only craft to operate as escort carriers. The arrival of the Sea Harrier VTOL/ STOVL fast jet meant they could carry fixed-wing aircraft, despite their short flight deck.

The U.S. used conventional carriers initially as pure ASW carriers, embarking helicopters and fixed-wing ASW aircraft like the S-2 Tracker. Later, specialized LPH helicopter carriers for the transport of United States Marine Corps troops and their helicopter transports were developed. These were evolved into the LHA and later into the LHD classes of amphibious assault ships, similar to the UK model even to the point of embarking Harrier Jump Jet aircraft, though much larger.

Ski-jump ramp

Still another British invention was the ski-jump ramp as an alternative to contemporary catapult systems. As the Royal Navy retired or sold the last of its World War II-era carriers, they were replaced with smaller ships designed to operate helicopters and the VTOL Sea Harrier fast jet; ships such as *HMS Invincible*. The ski-jump allowed Harriers to take off with heavier loads, a STOVL option allowing them to take off with a heavier payload despite its usage of space for aircraft parking. It has since been adopted by the navies of several nations.


India's light carrier INS *Viraat*, formerly HMS *Hermes*, purchased from the British, after INS Vikrant shows another example of the ski jump.

Post-World War II conflicts

UN carrier operations in the Korean War

The United Nations command began carrier operations against the North Korean Army on July 3, 1950 in response to the invasion of South Korea. Task Force 77 consisted at that time of the carriers USS *Valley Forge* and HMS *Triumph*. Before the armistice of July 27, 1953, 12 U.S. carriers served 27 tours in the Sea of Japan as part of the Task Force 77. During periods of intensive air operations as many as four carriers were on the line at the same time (see Attack on the Sui-ho Dam), but the norm was two on the line with a third "ready" carrier at Yokosuka able to respond to the Sea of Japan at short notice.

A second carrier unit, Task Force 95, served as a blockade force in the Yellow Sea off the west coast of North Korea. The task force consisted of a Commonwealth light carrier (HMS *Triumph*, *Theseus*, *Glory*, *Ocean*, and *HMAS Sydney)* and usually a U.S. escort carrier (USS *Badoeng Strait*, *Bairoko*, *Point*

Cruz, *Rendova*, and *Sicily*).

Over 301,000 carrier strikes were flown during the Korean War: 255,545 by the aircraft of Task Force 77; 25,400 by the Commonwealth aircraft of Task Force 95, and 20,375 by the escort carriers of Task Force 95. United States Navy and Marine Corps carrier-based combat losses were 541 aircraft. The Fleet Air Arm lost 86 aircraft in combat, and the Fleet Air Arm of Australia 15.

U.S. carrier operations in Southeast Asia

The United States Navy fought "the most protracted, bitter, and costly war" (René Francillon) in the history of naval aviation from August 2, 1964 to August 15, 1973 in the waters of the South China Sea. Operating from two deployment points (Yankee Station and Dixie Station), carrier aircraft supported combat operations in South Vietnam and conducted bombing operations in conjunction with the U.S. Air Force in North Vietnam under Operations Flaming Dart, Rolling Thunder, and Linebacker. The number of carriers on the line varied during differing points of the conflict, but as many as six operated at one time during Operation Linebacker.

Twenty-one aircraft carriers (all operational attack carriers during the era except *John F. Kennedy*) deployed to Task Force 77 of the U.S. Seventh Fleet, conducting 86 war cruises and operating 9,178 total days on the line in the Gulf of Tonkin. 530 aircraft were lost in combat and 329 more in operational accidents, causing the deaths of 377 naval aviators, with 64 others reported missing and 179 taken prisoner-of-war. 205 officers and men of the ship's complements of three carriers (*Forrestal*, *Enterprise*, and *Oriskany*) were killed in major shipboard fires.

Falklands War

During the Falklands War the United Kingdom was able to win a conflict 8,000 miles (13,000 km) from home in large part due to the use of the light fleet carrier HMS *Hermes* and the smaller "through deck cruiser" HMS *Invincible*. The Falklands showed the value of a VSTOL aircraft — the Hawker Siddeley Harrier (the RN Sea Harrier and press-ganged RAF Harriers) in defending the fleet and assault force from shore based aircraft and for attacking the enemy. Sea Harriers shot down 21 fast attack jets and suffered no aerial combat losses, although six were lost to accidents and ground fire. Helicopters from the carriers were used to deploy troops, medevac, SAR and ASW.

Operations in the Persian Gulf

The US has also made use of carriers in the Persian Gulf, Afghanistan and to protect its interests in the Pacific. During the 2003 invasion of Iraq US aircraft carriers served as the primary base of US air power. Even without the ability to place significant numbers of aircraft in Middle Eastern airbases, the United States was capable of carrying out significant air attacks from carrier-based squadrons. Recently, US aircraft carriers, such as the USS *Ronald Reagan* provided air support for counter-insurgency operations in Iraq.

Aircraft carriers today

Aircraft carriers are generally the largest ships operated by navies; a *Nimitz* class carrier powered by two nuclear reactors and four steam turbines is 1092 feet (333 m) long and costs about $4.5 billion. The United States has the majority of aircraft carriers (and also the only country with supercarriers) with eleven in service, one under construction, and one on order (it has to be noted that all of them are supercarriers). Its aircraft carriers are a cornerstone of American power projection capability.

Four modern aircraft carriers of various types – USS *John C. Stennis*, *FS Charles de Gaulle*, HMS *Ocean* and USS *John F. Kennedy* — and escort vessels on operations in 2002. The ships are sailing much closer together than they would during combat operations.

France's *Charles de Gaulle* (R-91), currently the only nuclear powered aircraft carrier operated by a country other than the United States.

Nine countries maintain a total of 21 aircraft carriers in active service: United States, United Kingdom, France, Russia, Italy, India, Spain, Brazil, and Thailand. In addition the People's Republic of China's People's Liberation Army Navy possesses the former Soviet aircraft carrier *Varyag*, but most naval analysts believe that they have no intention to operate it, but instead are using *Varyag* to learn about carrier operations for future Chinese aircraft carriers. The United States, South Korea, United Kingdom, Canada, the People's Republic of China, India, Japan, Australia, Chile, Singapore and France also operate vessels capable of carrying and operating multiple helicopters.

Aircraft carriers are generally accompanied by a number of other ships, to provide protection for the relatively unwieldy carrier, to carry supplies, and to provide additional offensive capabilities. This is often termed a battle group or carrier group, sometimes a carrier battle group.

In the early 21st century, worldwide aircraft carriers are capable of carrying about 1250 aircraft. US carriers account for over 1000 of these. The United Kingdom and France are both undergoing a major expansion in carrier capability (with a common ship class), but the United States will still maintain a very large lead.

Flight deck

As "runways at sea," modern aircraft carriers have a flat-top deck design that serves as a flight deck for take-off and landing of aircraft. Aircraft take off to the front, into the wind, and land from the rear. Carriers steam at speed, for example up to 35 knots (65 km/h), into the wind during take-off in order to increase the apparent wind speed, thereby reducing the speed of the aircraft relative to the ship. On some ships, a steam-powered catapult is used to propel the aircraft forward assisting the power of its engines and allowing it to take off in a shorter distance than would otherwise be required, even with the headwind effect of the ship's course. On other carriers, aircraft do not require assistance for take off — the requirement for assistance relates to aircraft design and performance. Conversely, when landing on a carrier, conventional aircraft rely upon a tailhook that catches on arrestor wires stretched across the deck to bring them to a stop in a shorter distance than normal. Other aircraft — helicopters and V/STOL (Vertical/Short Take-Off and Landing) designs — utilize their hover capability to land vertically and so require no assistance in speed reduction upon landing.

Conventional ("tailhook") aircraft rely upon a landing signal officer (LSO) to control the plane's landing approach, visually gauging altitude, attitude, and speed, and transmitting that data to the pilot. Before the angled deck emerged in the 1950s, LSOs used colored paddles to signal corrections to the pilot. From the late 1950s onward, visual landing aids such as mirrors provided information on proper glide slope, but LSOs still transmit voice calls to landing pilots by radio.

The flight deck of an aircraft carrier is one of the world's most dangerous places to work. To facilitate working on the flight deck of a U.S. aircraft carrier, the sailors wear colored shirts that designate their responsibilities. White shirts are responsible for safety. The LSO wears a white shirt. Red shirts handle munitions. Purple shirts (grapes) handle jet fuel. Yellow shirts are responsible for directing aircraft. Examples of yellow shirts are the shooter, the handler, and the air boss. The shooter, who is a pilot, is responsible for launching aircraft. The handler, who works just inside the island from the flight deck, is responsible for the movement of aircraft before launching and after landing. The air boss (usually a commander) occupies the top bridge and has the overall responsibility for controlling takeoffs, landings, "those aircraft in the air near the ship, and the movement of planes on the flight deck, which itself resembles a well-choreographed ballet." http://www.navy.mil/navydata/ships/carriers/powerhouse/powerhouse.asp The captain of the ship and the one star rear admiral of the carrier group do not wear colored shirts. The captain and his staff work in the command bridge below the top bridge. Below the command bridge is the flag bridge where the commander of the carrier group and his staff work.

Since the early 1950s it has been common to direct the landing recovery area off to port at an angle to the line of the ship. The primary function of the angled deck landing area is to allow aircraft who miss the arresting wires, referred to as a "bolter", to become airborne again without the risk of hitting aircraft parked on the forward parts of the deck. The angled deck also allows launching of aircraft at the same time as others land.

The above deck areas of the warship (the bridge, flight control tower, and so on) are concentrated to the starboard side of the deck in a relatively small area called an "island". The starboard side of the ship is used for the island because early carrier pilots showed a tendency to veer left in a crash situation. Very few carriers have been designed or built without an island and such a configuration has not been seen in a fleet sized carrier. The "flush deck" configuration proved to have very significant drawbacks, complicating navigation, air traffic control and numerous other factors.

F/A-18 Hornets on the flight deck of the *Nimitz*-class supercarrier *Harry S. Truman*

A more recent configuration, used by the British Royal Navy, has a 'ski-jump' ramp at the forward end of the flight deck. This was developed to help launch VTOL (or STOVL) aircraft (aircraft that are able to take off and land with little or no forward movement) such as the Sea Harrier. Although the aircraft are capable of flying vertically off the deck, using the ramp is more fuel efficient. As catapults and arrestor cables are unnecessary, carriers with this arrangement reduce weight, complexity, and space needed for equipment. The disadvantage of the ski jump — and hence, the reason this configuration has not appeared on American supercarriers — is the penalty that it exacts on aircraft size, payload and fuel load (and hence, range): Large, slow planes such as the E-2 Hawkeye and heavily-laden strike fighters like the F/A-18E/F Super Hornet cannot use a ski jump because their high weight requires either a longer takeoff roll than is possible on a carrier deck, or catapult assistance.

Future aircraft carriers

Several nations which currently possess aircraft carriers are in the process of planning new classes to replace current ones. The world's navies still generally see the aircraft carrier as the main future capital ship, with developments such as the arsenal ship, which have been promoted as an alternative, seen as too limited in terms of flexibility.

Military experts such as John Keegan have noted that in any future naval conflict between reasonably evenly matched powers, all surface ships - including aircraft carriers - would be at extreme and disproportionate risk, mainly due to the advanced capabilities of satellite reconnaissance and anti-ship missiles.

Contrary to the thrust of most current naval spending, Keegan therefore postulates that eventually, most navies will move to submarines as their main fighting ships, including in roles where submarines play only a minor or no role at the moment.

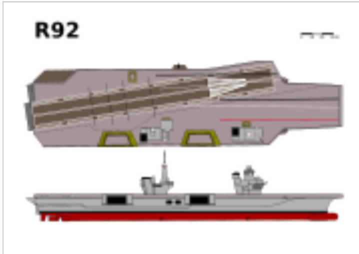
Chinese People's Liberation Army Navy

In June 2005, reports from boxun.com that the People's Republic of China would build a US$ 362 million aircraft carrier with a displacement of 78,000 tonnes were denied by Chinese defence official *Zhang Guangqin*.

China bought the unfinished Soviet aircraft carrier Varyag in 2001 from Ukraine, supposedly to be turned into a floating casino. Pictures taken while in port suggest this plan has been abandoned and show that work is being carried out to maintain its military function. There is no conclusive evidence as to what role it would play in the Chinese Navy.

In 2007, it was announced that China was working on a plan for producing its own aircraft carrier.

French Navy



Future French aircraft carrier

The French Navy has set in motion plans for a second CTOL aircraft carrier, to supplement *Charles de Gaulle*. The design is to be much larger, in the range of 65-74,000 tonnes, and will not be nuclear-powered like Charles de Gaulle. There are plans to buy the third carrier of the current Royal Navy design for CATOBAR operations (the Thales/BAE Systems design for the Royal Navy is for a STOVL carrier which is reconfigurable to CATOBAR operations).

Indian Navy

India started the construction of a 37,500 tonne, 252 meter-long *Vikrant* class aircraft carrier in April 2005. The new carrier will cost US$762 million and will operate MiG 29K 'Fulcrum', Naval HAL Tejas and Sea Harrier aircraft along with the Indian-made helicopter HAL Dhruv. The ship will be powered by four turbine engines and when completed will have a range of 7,500 nautical miles (14,000 km), carrying 160 officers, 1400 sailors, and 30 aircraft. The carrier is being constructed by a state-run shipyard in Cochin.

In 2004, India also bought *Admiral Gorshkov* from Russia for US$1.5 billion. It is most likely to be named the INS *Vikramaditya*, and is expected to join the Indian Navy in 2008 after a refit. However, this date now seems overly optimistic, as delays in the refit were announced in the middle of July 2007. Eduard Borisov, an acting director of Sevmash plant responsible for refit, stated that production capabilities of the plant were overestimated for current funding level, and the refit will be completed only in 2011. Vladimir Pastuhov, Sevmash director, had to step down, along with two other top managers of large defence contractors, in the largest scandal in the Russian defence industry in recent years.

Italian Navy

The construction of the conventional powered *Marina Militare* STOVL aircraft carrier *Cavour* began in 2001. It is being built by Fincantieri of Italy. After much delay, *Cavour* is expected to enter service in 2009 to complement the Marina Militare aircraft carrier *Giuseppe Garibaldi*.

Royal Navy (United Kingdom)

The Royal Navy is currently planning two new larger STOVL aircraft carriers (the *Queen Elizabeth* class) to replace the three *Invincible* class carriers. These two ships are to be named HMS *Queen Elizabeth* and HMS *Prince of Wales*. They will be able to operate up to 48 aircraft and will have a displacement of around 65,000 tonnes. The two ships are due to enter service in 2014 and 2016 respectively. Their primary aircraft complement will be made up of F-35B Lightning IIs, and their ship's company will number around 1000.

The two ships will be the largest warships ever built for the Royal Navy. Initially to be configured for STOVL operations, the carriers are to be adaptable to STOBAR or CATOBAR configurations to allow any type of future generation of aircraft to operate from them.

Russian Navy

Russian Navy Commander-in-Chief Admiral Vladimir Masorin officially stated on June 23, 2007, that Navy is currently considering a specifications of a new nuclear aircraft carrier design, for the class that was first announced about a month earlier. Production of the carriers is believed to start around 2010 at Zvezdochka plant in Severodvinsk, where the large drydock, capable of launching vessels with more than 100,000 ton displacement, is now being built.

In his statement Admiral Masorin stated that general dimensions of the project are already determined. The projected carrier is to have a nuclear propulsion, to displace about 50,000 tons and to carry an air wing of 30-50 air superiority aircraft and helicopters, which makes her roughly comparable to French Charles de Gaulle carrier. "The giants that the US Navy builds, those that carry 100-130 aircraft, we won't build anything like that", said Admiral Masorin. The planned specs reflects the role of aircraft carriers as an air support platforms for guided missile cruisers and submarines, traditional for the Russian Navy.

Russian naval establishment had long agreed that since the decommissioning of Kiev class carriers the only operational carrier Admiral Kuznetsov was insufficient, and that three or four carriers were necessary to meet the Navy's air support requirements. However financial and organisational turmoil of the 1990s made even maintenance of Admiral Kuznetsov a difficult undertaking. The recent improvement in Russia's economic situation has allowed a major increase in defence spending, and at least two new carriers were believed to be in planning, one each for Northern and Pacific fleets.

Spanish Navy

The project for the 231 meter-long, 25,000-30,000-tonne conventionally-powered *Juan Carlos I* for the Spanish Navy was approved in 2003, and its construction started in August 2005, with the shipbuilding firm Navantia in charge of the project. The *Juan Carlos I* is a vessel designed to operate both as amphibious assault ship vessel and as VSTOL aircraft carrier, depending on the mission assigned. The design was made keeping in mind the low-intensity conflicts in which the Spanish Navy is likely to be involved in the future. When it is configured to operate as VSTOL aircraft carrier, the operating range will be about 25,000 tonnes, and it will operate a maximum of 30 Matador AV-8B+, F-35 or a mixed force of both aircraft. The ship is provided with a Ski-Jump and a tri-dimensional radar based combat system, and she will be the second operating aircraft carrier of the Spanish navy after *Príncipe de Asturias*.

Australia is also purchasing two of these vessels, the *Canberra* class large amphibious ships, for their own navy.

US Navy

The current US Fleet of *Nimitz* class carriers are to be followed into service (and in some cases replaced) by the *Gerald R. Ford* (CVN-78) class. It is expected that the ships will be larger than the *Nimitz*, and will also be designed to be less detectable by radar. The United States Navy is also looking to make these new carriers more automated in an effort to reduce the amount of funding required to maintain and operate its supercarriers.

Image:CVN-21.jpg
Virtual depiction of the new US Navy *Gerald R. Ford-Class* carrier

With the decommissioning of the *USS John F. Kennedy* in March, 2007, the US fleet has been reduced to 11 supercarriers; thus creating major discussions between the Joint Chiefs of Staff and Congress. The House Armed Services Seapower subcommittee on July 24, 2007 is recommending 7, maybe 8 new carriers (1 every 4 years). However, the debate is deepened over budgeting for the $12-14.5 billion (plus $12 billion for development and research) for the *Gerald Ford*-class carrier (estimated service 2015). And, comparing these expenditures for a smaller $2 billion 45,000-ton class big-deck amphibious assault ships for squadrons of the new F-35Bs.

Airship

2008/9 Schools Wikipedia Selection. Related subjects: Air & Sea transport

USS Akron (ZRS-4) in flight, November 2, 1931

An **airship** or **dirigible** is a lighter than air (buoyant) aircraft that can be steered and propelled through the air using rudders and propellers. Unlike other *aerodynamic* aircraft such as fixed-wing aircraft (airplanes) and helicopters, which produce lift by moving a wing or airfoil through the air, *aerostatic* aircraft, such as airships and hot air balloons, stay aloft by filling a large cavity, such as a balloon, with a lighter than air gas.

The main types of airship are Non-rigid airships (or blimps), semi-rigid airships and rigid airships. Blimps are small airships without internal skeletons. Semi-rigid airships are slightly larger and have some form of internal support such as a fixed keel. Rigid airships with a full skeleton, such as the massive Zeppelin transoceanic models, are now a thing of the past.

Airships were the first aircraft to make controlled, powered flight. They were widely used before the 1940s. Their use decreased over time as their capabilities were surpassed by those of airplanes. Their decline furthered with a series of high-profile accidents, including the 1937 burning of the hydrogen-filled *Hindenburg* near Lakehurst, New Jersey. Airships are still used today in certain niche applications, such as advertising and as a camera platform for sporting events.

Terminology

In many countries, airships are also known as *dirigibles* from the French (*dirigir* to direct plus -ible), meaning "directable" or steerable. The first airships were called *dirigible balloons.* Over time, the word *balloon* was dropped from the phrase. In the modern usage, balloon refers to buoyant aircraft that generally rely on wind currents for movement, though vertical movement can be controlled in both.

The term zeppelin is a genericised trademark that originally referred to airships manufactured by the Zeppelin Company. Their crafts' names were usually prefixed with the word *Luftschiff*, German for "airship".

In modern common usage, the terms *zeppelin*, *dirigible* and *airship* are used interchangeably for any type of rigid airship, with the terms *blimp* or *airship* alone used to describe non-rigid airships. Although the blimp also qualifies as a "dirigible", the term is seldom used with blimps. In modern technical usage, *airship* is the term used for all aircraft of this type, with *zeppelin* referring only to aircraft of that manufacture, and *blimp* referring only to non-rigid airships.

The term *airship* is sometimes informally used to mean any machine capable of atmospheric flight.

There is often some confusion around the term *aerostat* with regard to airships. This confusion arises because *aerostat* has two different meanings. One meaning of *aerostat* refers to all craft that remain aloft using buoyancy. In this sense, airships are a type of *aerostat*. The other, more narrow and technical meaning of *aerostat* refers only to tethered or moored balloons. In this second technical sense, airships are distinct from *aerostats*. This airship/aerostat confusion is often exacerbated by the fact that both airships and aerostats have roughly similar shapes and comparable tail fin configurations, although only airships have motors.

Types

- Non-rigid airships (blimps) use a pressure level in excess of the surrounding air pressure in order to retain their shape.
- Semi-rigid airships, like blimps, require internal pressure to maintain their shape, but have extended, usually articulated keel frames running along the bottom of the envelope to distribute suspension loads into the envelope and allow lower envelope pressures.
- Rigid airships (Zeppelin is almost synonymous with this type) have rigid frames containing multiple, non-pressurized gas cells or balloons to provide lift. Rigid airships do not depend on internal pressure to maintain their shape and can be made to virtually any size.
- Metal-clad airships had characteristics of both rigid and non-rigid airships, utilizing a very thin, airtight metal envelope, rather than the usual rubber-coated fabric envelope. Only four ships of this type, Schwarz's aluminium ships of 1893 and 189 the ZMC-2 and the Slate "City of Glendale", have been built to date with only the ZMC-2 a success.
- Hybrid airship is a general term for an aircraft that combines characteristics of heavier-than-air (airplane or helicopter) and lighter than air technology. Examples include helicopter/airship hybrids intended for heavy lift applications and dynamic lift airships intended for long-range cruising. It should be noted that most airships, when fully loaded with cargo and fuel, are typically heavier than air, and thus must use their propulsion system and shape to generate aerodynamic lift, necessary to stay aloft; technically making them hybrid airships. However, the term "hybrid airship" refers to craft that obtain a significant portion of their lift from aerodynamic lift and often require substantial take-off rolls before becoming airborne.

Image:Airship types.gif

Rigid, Semi-rigid and Non-rigid airship types

In the background, ZR-3, in front of it, (l to r) J-3 or 4, K-1, ZMC-2, in front of them, "Caquot" observation balloon, and in foreground free balloons used for training. US Navy airships and balloons, 1931

Lifting gas

Any gas that is lighter than air can be used to create buoyant lift, however many such gases are either toxic, flammable, corrosive, or a combination of these, limiting their use in airships. Historically, hydrogen and helium have been used in large airships. A calculation based on the gas densities shows that hydrogen provides only 8% more lift than helium.

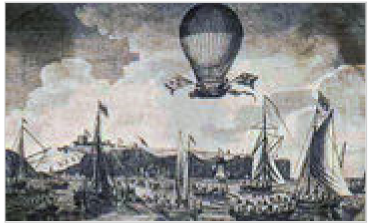
After the discovery of helium in the late 1890s, and development of processes to produce the gas commercially, helium was the preferred lifting gas. However until the 1950s, the United States was the sole producer of helium, and because the U.S. had embargoed exports of helium to Germany for strategic purposes since the 1920s, German airships were filled with hydrogen. The *Hindenburg*, for example, was originally designed to be filled with helium, but its unavailability forced the airship's operators to use hydrogen, with infamous results.

Ships called thermal airships utilize heated air, in a fashion similar to hot air balloons, as their lifting gas.

History

Early pioneers

In 1784 Jean-Pierre Blanchard fitted a hand-powered propeller to a balloon, the first recorded means of propulsion carried aloft. In 1785, he crossed the English Channel with a balloon equipped with flapping wings for propulsion, and a bird-like tail for steerage.



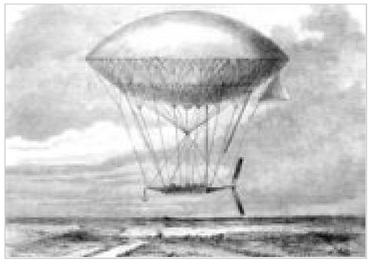
Crossing of the English Channel by Blanchard in 1785.

The first person to make an engine-powered flight was Henri Giffard who, in 1852, flew 27 km (17 miles) in a steam-powered airship. Airships would develop considerably over the next two decades: In 1863, Dr. Solomon Andrews devised the first fully steerable airship, although it had no motor. In 1872, the French naval architect Dupuy de Lome launched a large limited navigable balloon, which was driven by a large propeller and the power of eight people. It was developed during the Franco-Prussian war, as an improvement to the balloons used for communications between Paris and the countryside during the Siege of Paris by German forces, but was only completed after the end of the war. Charles F. Ritchel made a public demonstration flight in 1878 of his hand-powered one-man rigid airship and went on to build and sell five of his aircraft. Paul Haenlein flew an airship with an internal combustion engine running on the coal gas used to inflate the envelope over Vienna, the first use of such an engine to power an aircraft in 1872.

A model of the Giffard Airship at the London Science Museum.

In the 1880s a Serb named Ogneslav Kostovic Stepanovic also designed and built an airship. However the craft was destroyed by fire before it flew. In 1883, the first electric-powered flight was made by Gaston Tissandier who fitted a 1.5 hp (1 kW) Siemens electric motor to an airship. The first fully controllable free-flight was made in a French Army airship, *La France*, by Charles Renard and Arthur Constantin Krebs in 1884 . The 170-foot (52 m) long, 66,000 cubic foot (1,900 m^3) airship covered 8 km (5 miles) in 23 minutes with the aid of an 8.5 hp (6 kW) electric motor.

In 1888-97, Dr. Frederich Wölfert built three Daimler Motor Company-built petrol engine powered airships, the last of which caught fire in flight and killed both occupants.



The navigable balloon developed by Dupuy de Lome in 1872.

Santos-Dumont #6 rounding the Eiffel Tower, winning the Deutsch Prize in 1901.

In 1896, a rigid airship created by Croatian engineer David Schwarz made its first flight at Tempelhof field in Berlin. After Schwarz's death, his wife, Melanie Schwarz, was paid 15,000 Marks by Count Ferdinand von Zeppelin for information about the airship.

In 1901, Alberto Santos-Dumont, in his airship "Number 6", a small blimp, won the Deutsch de la Meurthe prize of 100,000 francs for flying from the Parc Saint Cloud to the Eiffel Tower and back in under thirty minutes. Many inventors were inspired by Santos-Dumont's small airships and a veritable airship craze began world-wide. Many airship pioneers, such as the American Thomas Scott Baldwin financed their activities through passenger flights and public demonstration flights. Others, such as Walter Wellman and Melvin Vaniman set their sights on loftier goals, attempting two polar flights in 1907 and 1909, and two trans-atlantic flights in 1910 and 1912.

"The Golden Age"

The "Golden Age of Airships" began in July 1900 with the launch of the Luftschiff Zeppelin LZ1. This led to the most successful airships of all time: The Zeppelins. These were named after Count von Zeppelin who began experimenting with rigid airship designs in the 1890s leading to the badly flawed LZ1 (1900) and the more successful LZ2 (1906). At the beginning of World War I the Zeppelin airships had a framework composed of triangular lattice girders, covered with fabric and containing separate gas cells. Multi-plane, later cruciform, tail fins were used for control and stability, and two engine/crew cars hung beneath the hull driving propellers attached to the sides of the frame by means of long drive shafts. Additionally there was a passenger compartment (later a bomb bay) located halfway between the two cars.

First World War

The prospect of airships as bombers had been recognised in Europe well before the airships were up to the task. H. G. Wells *The War in the Air* (1908) described the obliteration of entire fleets and cities by airship attack. On 5 March 1912, Italian forces became the first to use dirigibles for a military purpose during reconnaissance west of Tripoli behind Turkish lines. It was World War I, however, that marked the airship's real debut as a weapon.

Caquot observation dirigeable during the First World War.

Albert Caquot designed an Observation Balloon for the French army in 1914. The Type R Observation balloon was used by all the allied forces, including the British and United States Armies, at the end of the World War. In 1919, Japan equipped the Imperial Army with several "Caquot dirigeables".

The Germans, French and Italians all operated airships in scouting and tactical bombing roles early in the war, and all learned that the airship was too vulnerable for operations over the front. The decision to end operations in direct support of armies was made by all in 1917. Count Zeppelin and others in the German military believed they had found the ideal weapon with which to counteract British Naval superiority and strike at Britain itself. More realistic airship advocates believed the Zeppelin was a valuable long range scout/attack craft for naval operations. Raids began by the end of 1914, reached a first peak in 1915, and then were discontinued in August 1918. Zeppelins proved to be terrifying but inaccurate weapons. Navigation, target selection and bomb-aiming proved to be difficult under the best of conditions. The darkness, high altitudes and clouds that were frequently encountered by zeppelin missions reduced accuracy even further. The physical damage done by the zeppelins over the course of the war was trivial, and the deaths that they caused (though visible) amounted to a few hundred at most. The zeppelins also proved to be vulnerable to attack by aircraft and antiaircraft guns, especially those armed with incendiary bullets. Several were shot down in flames by British defenders, and others crashed 'en route'. In retrospect, advocates of the naval scouting role of the airship proved to be correct, and the land bombing campaign proved to be disastrous in terms of morale, men and material. Many pioneers of the German airship service died in what was the first strategic bombing campaign in history. Countermeasures by the British were sound detection, equipment, search lights and anti-aircraft artillery, and starting in 1915 night fighters. One method used early in the war when short range meant the airships had to fly from forward bases, and when only Zeppelin production facilities were in Friedrichshafen, was bombing of airship sheds by the British Royal Naval Air Service. Late in the war, the development of the aircraft carrier led to the first successful carrier air strike in history. The morning of 19 July 1918, seven Sopwith 2F.1 Camels were launched from HMS Furious and struck the airship base at Tondern, destroying the Zeppelins L 54 and L 60.


View from a French dirigible approaching a ship in 1918.

Before the World War, the British Army was interested in blimps for scouting purposes. The Royal Navy recognizing the potential threat that scouting Zeppelins might pose, decided in 1908 to produce an example of rigid airship so that the threat might be evaluated in practice instead of theory. The Royal Navy was to continue development of rigid airships until the end of the war. The British Army abandoned airship development in favour of airplanes by the start of the war, but the Royal Navy had recognised the need for small airships to counteract the submarine and mine threat in coastal waters. Beginning in February 1915, began to deploy the SS (Sea Scout) class of blimp. These had a small envelope of 60-70,000 cu feet and at first utilised standard single engined planes (BE2c, Maurice Farman, Armstrong FK) shorn of wing and tail surfaces as control cars, an economy measure. Eventually more advanced blimps with purpose built cars, such as the C (Coastal), C* (Coastal Star), NS (North Sea), SSP (Sea Scout Pusher), SSZ (Sea Scout Zero), SSE (Sea Scout Experimental) and SST (Sea Scout Twin) classes were developed. The NS class, after initial teething problems proved to be the largest and finest airships in British service. They had a gas capacity of 360,000 cu feet, a crew of 10 and an endurance of 24 hours. Six 230 lb bombs were carried, as well as 3-5 machine guns. British blimps were used for scouting, mine clearance, and submarine attack duties. During the war, the British operated 226 airships, mostly non-rigid, most of which were of indigenous construction, though some non-rigid airships operated were purchased from France and even Germany (before the war). Of that number several were sold to Russia, France, the US and Italy. Britain, in turn, purchased one M-type semi-rigid from Italy whose delivery was delayed until 1918. Nine rigid airships had been completed by the armistice, although several more were in an advanced state of completion by the war's end. The large number of trained crews, low attrition rate and constant experimentation in handling techniques meant that at the war's end Britain was the world leader in non-rigid airship technology.

Both France and Italy continued airships throughout the war. France preferred non-rigid types while Italy operated 49 semi-rigid airships in both the scouting and bombing roles.

Airplanes had essentially replaced airships as bombers by the end of the war, and Germany's remaining zeppelins were scuttled by their crews, scrapped or handed over to the Allied powers as spoils of war. The British rigid airship program, meanwhile, had been largely a reaction to the potential threat of the German one and was largely, though not entirely, based on imitations of the German ships.

Inter-war period

Airships were operated in a number of nations between the two world wars. The major operators of rigid airships were Britain, the United States and Germany, and a few were operated by Italy and France. Italy, the Soviet Union, United States and Japan operated semi-rigid airships, while blimps were operated in many nations.

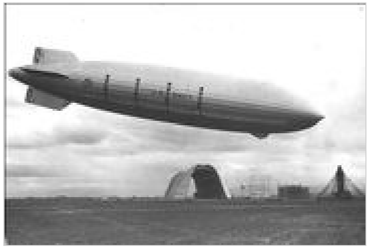
The British R33 and R34 were near identical copies of the German L 33, which crashed virtually intact in Yorkshire on September 24, 1916. Despite being almost three years out of date by the time they were launched in 1919, they were two of the most successful in British service. The creation of the Royal Air Force (RAF) in early 1918 created a hybrid British airship program. The RAF was uninterested in airships and the Admiralty was, so a deal was made where the Admiralty would design any future military airships while the RAF would handle manpower, facilities and operations.

Rescuers scramble across the wreckage of British R-38/USN ZR-2, August 24, 1921

After the armistice, the airship program was rapidly wound down, and rigid airship operations were curtailed. On July 2, 1919 R34 began the first double crossing of the Atlantic by an aircraft. It landed at Mineola, Long Island on July 6, 1919 after 108 hours in the air. The return crossing commenced on July 8 because of concerns about mooring the ship in the open, and took 75 hours. Impressed, British leaders began to contemplate a fleet of airships to link Britain to its far-flung colonies. But post-war economic conditions led to scrapping most airships and dispersion of trained personnel, until starting construction of the R-100 and R-101 in 1929. The major consequence of Britain's interest in establishing airship service to the empire was the effort to use the Allies' seizure of German airships and airship sheds to avoid competition from Germany. The US Navy contracted to buy the British built R-38, but before that airship was turned over to the US, it was lost to structural failure due to both improper design and operation.

The first American-built rigid dirigible was USS *Shenandoah*, christened on August 20 in Lakehurst, New Jersey. It flew in 1923, while the *Los Angeles* was under construction. It was the first ship to be inflated with the noble gas helium, which was still so rare that the *Shenandoah* contained most of the world's reserves. When the *Los Angeles* was delivered, the two airships had to share the limited supply of Helium, and thus alternated operating and overhauls.

Construction of the USS Shenandoah (ZR-1), 1923



US Navy Zeppelin ZRS-5 "USS Macon" over Moffett Field in 1933

The United States Navy purchased what became the USS *Los Angeles* and paid with "war reparations" money, owed according to the Versailles Treaty, thus saving The Zeppelin works. The success of the *Los Angeles* encouraged the US Navy to invest in its own, larger airships. The USS *Los Angeles* flew successfully for 8 years.

Meanwhile Germany was building the *Graf Zeppelin*, the largest airship that could be built in the company's existing shed, and intended to stimulate interest in passenger airships. The *Graf Zeppelin* burned *blau gas*, similar to propane, stored in large gas bags below the hydrogen cells, as fuel. Since its density was similar to that of air, it avoided the weight change when fuel was used, and thus the need to valve hydrogen. The "Graf" was a great success and compiled an impressive safety record. For example it flew over one million miles (including the first circumnavigation of the globe by air) without a single passenger injury.

The US Navy developed the idea of using airships as "flying aircraft carriers." There were two airships, the world's largest at the time, to test the principle—the USS *Akron* and USS *Macon*. Each carried four fighters in their "hanger", and could carry a fifth on the "trapeze." The "Flying Aircraft Carrier" had mixed results. By the time the Navy started to develop a sound doctrine for using the ZRS-type airships, the last of the two built, USS *Macon*, was lost. The seaplane had become more mature, and was considered a better investment.

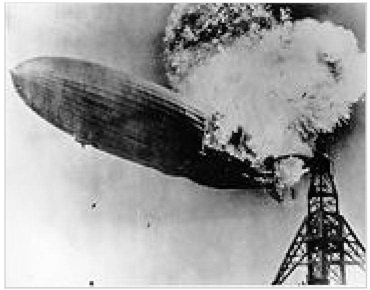

The USS Akron over Manhattan circa 1932

Eventually the US Navy lost all three American-built rigid airships to accidents. USS *Shenandoah* on a poorly planned publicity flight flew into a severe thunderstorm over Noble County, Ohio on 3 September 1925. It broke into pieces, killing 14 of her crew. USS *Akron* was caught in a severe storm and flown into the surface of the sea off the shore of New Jersey on April 3, 1933. It carried no life boats and few life vests, so 73 of her 76-men crew died from drowning or hypothermia. USS *Macon* was lost after suffering a structural failure off the shore of Point Sur, California on 12 February 1935. The failure caused a loss of gas, which was made much worse when the aircraft was driven over pressure height causing it to lose too much helium to maintain flight. Only 2 of her 83-man crew died in the crash thanks to the inclusion of life jackets and inflatable rafts after the *Akron* disaster.

Britain's Burney Scheme and decline in airships

In Britain during the 1920s, Sir Denistoun Burney suggested a plan for air service throughout the Empire by airships (the Burney Scheme). Following the election of Ramsey MacDonald, the Burney scheme was transformed into a government-controlled program which contracted for two airships, one to be developed by the Airship Guarantee Company, the other by the Royal Airship Works. The two designs were radically different. The "capitalist" ship, the *R100*, was conservative, while the "socialist" ship, the *R101*, was wildly innovative. Construction was delayed, and the airships did not fly until 1929. Neither airship was capable of the service intended, though the R100 did complete a proving flight to Canada and back in 1930.

In October 1930 there were rushed preparations to fly the *R-101*, which had not been adequately tested and had serious deficiencies, on a flight to India carrying the Air Minister of the MacDonald government, Christopher Birdwell, Lord Thompson for an important Imperial conference. An air worthiness certificate was issued at the last moment. The R101 left on the flight on 5 October but hours later crashed in France killing 48 of the 54 people aboard. Because of the bad publicity surrounding the crash, the Air Ministry grounded the competing R100 in 1930 and sold it for scrap in 1931, ending the era of British rigid airships.


The *Hindenburg* — moments after catching fire, May 6, 1937

By the mid-1930s only Germany still pursued the airship. The Zeppelin company continued to operate the *Graf Zeppelin* on passenger service between Germany and Brazil. Even with the small *Graf Zeppelin*, the operation was almost profitable. In the mid-1930s work started to build an airship designed specifically to operate a passenger service across the Atlantic. The *Hindenburg* completed a very successful 1936 season carrying passengers between Lakehurst, New Jersey and Germany. But 1937 started with the most spectacular and widely remembered airship accident. Approaching the mooring mast minutes before landing on 6 May 1937, the *Hindenburg* burst into flames and crashed. Of the 97 people aboard, 36 died: 13 passengers, 22 aircrew, and one American ground-crewman. The disaster happened before a large crowd, was filmed and a radio news reporter was cutting a recording of his coverage of the arrival. This was a disaster which theatre goers could see and hear the next day. On that same next day, the *Graf Zeppelin* landed at the end of its flight from Brazil, ending intercontinental passenger airship travel.

Hindenburg's sister ship, the *Graf Zeppelin II*, could not fly without helium which the United States refused to sell. The *Graf Zeppelin* flew some test flights and conducted electronic espionage until 1939 when it was grounded due to the start of the war. The last two Zeppelins were scrapped in 1940.

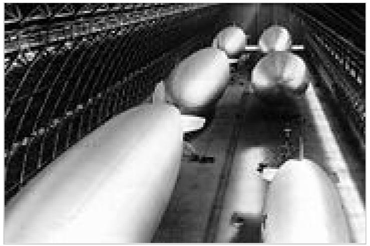
Development of airships continued only in the United States, and in a small way, the Soviet Union.

Second World War

While Germany determined that airships were obsolete for military purposes in the coming war and concentrated on the development of airplanes, the United States pursued a program of military airship construction even though it had not developed a clear military doctrine for airship use. At the Japanese attack on Pearl Harbour on 7 December 1941 that brought the United States into World War II, it had 10 non-rigid airships:

- 4 K-class: K-2, K-3, K-4 and K-5 designed as a patrol ships built from 1938.
- 3 L-class: L-1, L-2 and L-3 as small training ships, produced from 1938.
- 1 G-class built in 1936 for training.
- 2 TC-class that were older patrol ships designed for land forces, built in 1933. The US Navy acquired them from Army in 1938.

Only K and TC class airships were suitable for combat and they were quickly pressed into service against Japanese and German submarines which were then were sinking US shipping within visual range of the US coast. US Navy command, remembering the airship anti-submarine success from World War I, immediately requested new modern anti-submarine airships and on 2 January 1942 formed the ZP-12 patrol unit based in Lakehurst from the 4 K airships. The ZP-32 patrol unit was formed from two TC and two L airships a month later, based at NAS Moffett Field in Sunnyvale, California. An airship training base was created there as well. In December 1941 and the first months of 1942, the Goodyear blimp *Resolute* was operated as an anti-submarine privateer based out of Los Angeles. As the only US craft to operate under a Letter of Marque since the War of 1812, the *Resolute*, armed with a rifle and flown by its civilian crew, patrolled the seas for submarines.



A view of six helium-filled blimps being stored in one of the two massive hangars located at NAS Santa Ana, during World War II.

In the years 1942–44, approximately 1,400 airship pilots and 3,000 support crew members were trained in the military airship crew training program and the airship military personnel grew from 430 to 12,400. The US airships were produced by the Goodyear factory in Akron, Ohio. From 1942 till 1945, 154 airships were built for the US Navy (133 K-class, ten L-class, seven G-class, four M-class) and five L-class for civilian customers (serial number L-4 to L-8).

The primary airship tasks were patrol and convoy escort near the US coastline. They also served as an organisation centre for the convoys to direct ship movements, and were used in naval search and rescue operations. Rarer duties of the airships included aerophoto reconnaissance, naval mine-laying and mine-sweeping, parachute unit transport and deployment, cargo and personnel transportation. They were deemed quite successful in their duties with the highest combat readiness factor in the entire US air force (87%).

During the war some 532 ships without airship escort were sunk near the US coast by enemy submarines. Only one ship, the tanker Persephone, of the 89,000 or so in convoys escorted by blimps was sunk by the enemy. Airships engaged submarines with depth charges and, less frequently, with other on-board weapons. They were excellent at driving submarines down, where their limited speed and range prevented them from attacking convoys. The weapons available to airships were so limited that until the advent of the homing torpedo they had little chance of sinking a submarine.

Only one airship was ever destroyed by U-boat: on the night of 18/ 19 July 1943 a K-class airship (K-74) from ZP-21 division was patrolling the coastline near Florida. Using radar, the airship located a surfaced German submarine. The K-74 made her attack run but the U-boat opened fire first. K-74's depth charges did not release as she crossed the U-boat and the K-74 received serious damage, losing gas pressure and an engine but landing in the water without loss of life. The crew was rescued by patrol boats in the morning, but one crewman, Isadore Stessel, died from a shark attack. The U-Boat, U-134, was slightly damaged and the next day or so was attacked by aircraft sustaining damage that forced it to return to base. It was finally sunk on 24 August 1943 by a British Vickers Wellington near Vigo, Spain

Fleet Airship Wing One operated from Lakehurst, NJ, Glynco, GA, Weeksville, NC, South Weymouth NAS Massachusetts, Brunswick NAS and Bar Harbour ME, Yarmouth, Nova Scotia, and Argentia, Newfoundland.

Some US airships saw action in the European war theatre. The ZP-14 unit operating in the Mediterranean area from June 1944 completely denied the use of the Gibraltar Straits to Axis submarines. Airships from the ZP-12 unit took part in the sinking of the last U-Boat before German capitulation, sinking U-881 on 6 May 1945 together with destroyers Atherton and Mobery.

Other airships patrolled the Caribbean, Fleet Airship Wing Two, Headquartered at NAS Richmond, Florida, covered the Gulf of Mexico from Richmond and Key West, FL, Houma, Louisiana, as well as Hitchcock and Brownsville, Texas. FAW 2 also patrolled the northern Caribbean from San Julian, the Isle of Pines and Guantanamo Bay, Cuba as well as Vernam Field Jamaica.

Navy blimps of Fleet Airship Wing Five, (ZP-51) operated from bases in Trinidad, British Guiana and Parmaribo, Dutch Guiana. Fleet Airship Wing Four operated along the coast of Brazil. Two squadrons, VP-41 and VP-42 flew from bases at Amapá, Igarape Assu, Sao Luiz, Fortaleza, Fernando de Noronha, Recife, Maceiro, Ipitanga, Caravellas, Vitoria and the hanger built for the *Graf Zeppelin* at Santa Cruz.

Fleet Airship Wing Three operated squadrons, ZP-32 from Moffett Field, ZP-31 at NAS Santa Anna, and ZP-33 at Tillamook Oregon. Auxiliary fields were at Del Mar, Lompoc, Watsonville and Eureka, CA, North Bend and Astoria, Oregon, as well as Shelton and Quillayute in Washington.

From 2 January 1942 till the end of war airship operations in the Atlantic, the airships of the Atlantic fleet made 37,554 flights and flew 378,237 hours. Of the over 70,000 ships in convoys protected by blimps, only one was sunk by a submarine while under blimp escort.

The Soviet Union used a single airship during the war. The W-12, built in 1939, entered service in 1942 for paratrooper training and equipment transport. It made 1432 runs with 300 metric tons of cargo until 1945. On 1 February 1945 the Soviets constructed a second airship, a Pobieda-class (*Victory*-class) unit (used for mine-sweeping and wreckage clearing in the Black Sea) which crashed on 21 January 1947. Another W-class — W-12bis Patriot was commissioned in 1947 and was mostly used for crew training, parades and propaganda.

Modern use

One of The Goodyear Tire & Rubber Company's blimp fleet.

Although airships are no longer used for passenger transportation, they are still used for other purposes such as advertising and sightseeing.

In June, 1987, the US Navy awarded a US$168.9 million contract to Westinghouse Electric and Airship Industries of the UK to demonstrate whether a blimp could be used as an airborne platform to detect the threat of sea-skimming missiles, such as the Exocet.

In recent years, the Zeppelin company has reentered the airship business. Their new model, designated the Zeppelin NT made its maiden flight on September 18, 1997. There are currently three NT aircraft flying. One was sold to a Japanese company, and was planned to be flown to Japan in the summer of 2004. But due to delays getting permission from the Russian government, the company decided to transport the airship to Japan by ship.

Blimps are used for advertising and as TV camera platforms at major sporting events. The most iconic of these is the Goodyear blimps. Goodyear operates 3 blimps in the United States, and the Lightship group operates up to 19 advertising blimps around the world. Airship Management Services owns and operates 3 Skyship 600 blimps. Two operate as advertising and security ships in the North America and the Caribbean.

Skycruise Switzerland AG owns and operates 2 skyship 600 blimps. One operates regularly over Switzerland used on sightseeing tours.

The Switzerland-based Skyship 600 has also played other roles over the years. For example, it was also flown over Athens during the 2004 Summer Olympics as a security measure. In November 2006, it carried advertising calling it "The Spirit of Dubai" as it began a publicity tour from London to Dubai, UAE on behalf of The Palm Islands, the worlds largest man-made islands created as a residential complex.

Los Angeles-based Worldwide Aeros Corp. produces FAA Type Certified Aeros 40D Sky Dragon airships.


The Spirit of Dubai approaches its motorised mooring mast.

In May 2006, US Navy began to fly airships again after a hiatus of nearly 44 years. The program uses a single American Blimp Company A-170 non-rigid airship. Operations focus on crew training and research, and the platform integrator is Northrop Grumman. The program is directed by the Naval Air Systems Command and is being carried out at NAES Lakehurst, the original centre of US Navy lighter-than-air operations in previous decades.

In November 2006, the US Army bought an A380+ airship from American Blimp Corporation through a Systems level contract with Northrop Grumman and Booz Allen Hamilton. The airship will start flight tests in late 2007 with a primary goal of carrying 2,500 lb of payload to an altitude of 15,000ft under remote control and autonomous waypoint navigation. The program will also demonstrate carrying 1,000 lb of payload to 20,000ft. The platform could be used for Multi-Intelligence collections. Northrop Grumman (formerly Westinghouse) has responsibility for the overall program.

Hot air airship

Several companies, such as Cameron Balloons in Bristol, United Kingdom, build hot-air airships. These combine the structures of both hot-air balloons and small airships. The envelope is the normal 'cigar' shape, complete with tail fins, but is inflated with hot air (as in a balloon) to provide the lifting force, instead of helium. A small gondola, carrying the pilot (and sometimes between 1 and 3 passengers), a small engine and the burners to provide the hot air is suspended below the envelope, below an opening through which the burners protrude.

Hot-air airships typically cost less to buy and maintain than modern helium-based blimps, and can be quickly deflated after flights. This makes them easy to carry in trailers or trucks and inexpensive to store. They are usually very slow moving, with a typical top speed of 15-20 mph. They are mainly used for advertising, but at least one has been used in rainforests for wildlife observation, as they can be easily transported to remote areas.

Remote controlled (RC) airships, a type of Unmanned Aerial System (UAS), are sometimes used for commercial purposes such as advertising and aerial video and photography as well as recreational purposes. They are particularly common as an advertising mechanism at indoor stadiums. While RC airships are sometimes flown outdoors, doing so for commercial purposes is illegal in the US In particular, Docket FAA-2006-25714 states that: "The FAA recognizes that people and companies other than modelers might be flying UAS with the mistaken understanding that they are legally operating under the authority of AC 91-57. AC 91-57 only applies to modelers, and thus specifically excludes its use by persons or companies for business purposes."

Present-day research

Prototypes and experimental models

Hybrid designs such as the Heli-Stat airship/helicopter, the Aereon aerostatic/aerodynamic craft, and the Cyclocrane (a hybrid aerostatic/rotorcraft), have struggled to take flight. The Cyclocrane was also interesting in that the airship's envelope rotated along its longitudinal axis.

CL160 was a very large semi-rigid airship to be built in Germany by the start-up Cargolifter, but funding ran out in 2002 after a massive hangar was built. The hangar, built just outside Berlin, has since been converted into a resort called " Tropical Islands".

In 2005, a short-lived project of the US Defense Advanced Research Projects Agency (DARPA) was WALRUS HULA which explored the potential for using airships as long-distance, heavy lift craft. The primary goal of the research program was to determine the feasibility of building an airship capable of carrying 500 short tons (450 metric tons) of payload a distance of 12,000 miles (20,000 km) and land on an unimproved location without the use of external ballast or ground equipment (such as masts). In 2005, two contractors, Lockheed-Martin and US Aeros Airships were each awarded approximately $3 million to do feasibility studies of designs for WALRUS. In late March 2006, DARPA announced the termination of work on WALRUS after completion of the current Phase I contracts.

The US government is funding two major projects in the high altitude arena. The Composite Hull High Altitude Powered Platform (CHHAPP) is sponsored by US Army Space and Missile Defense Command. This aircraft is also sometimes called *HiSentinel High-Altitude Airship*. This prototype ship made a 5-hour test flight in September 2005. The second project, the high-altitude airship (HAA), is sponsored by DARPA. In 2005, DARPA awarded a contract for nearly $150 million to Lockheed-Martin for prototype development. First flight of the HAA is planned for 2008.

Many companies are working on high-altitude airships. Techsphere is developing a high-altitude version of their spherically shaped airships. JP Aerospace has discussed its long-range plans that include not only high altitude communications and sensor applications but also an " orbital airship" capable of lifting cargo into low Earth orbit with a marginal transportation cost of $1 per short ton per mile of altitude.

On January 31, 2006 Lockheed-Martin made the first flight of their secretly built hybrid-airship designated the P-791 at the company's flight test facility on the Palmdale Air Force Plant 42. The design is very similar in to the SkyCat, unsuccessfully promoted for many years by the now financially troubled British company Advanced Technology Group. Although Lockheed Martin is developing a design for the DARPA WALRUS HULA project, it claimed that the P-791 is unrelated to WALRUS. Nonetheless, the design represents an approach that may well be applicable to WALRUS. Some believe that Lockheed-Martin had used the secret P-791 program as a way to get a "head-start" on the other WALRUS competitor, Aeros.

A privately funded effort to build a heavy-lift aerostatic/aerodynamic hybrid craft, called the Dynalifter, is being carried out by Ohio Airships. Test flights are to begin in Spring 2006.

The research and development company for airship technologies, 21st century Airships Inc., has developed a spherical-shaped airship, and airships for high altitude, environmental research, surveillance and military applications, heavy lift and sightseeing. Its airships have set numerous world records.

In Russia AUGUR-RosAerosystems Group is manufacturing non-rigid multi-functional airships for up to 10 passengers, as well as patrol airships including the Au-12 and Au-30. They are also working on developmental programs for heavy-lift cargo models and high-altitude stratospheric ships. One of AUGUR-RosAeroSystems manufactured Au-30 airship will take part in the expedition to the North Pole challenged by famous French explorer Jean-Louis Etienne for arctic ice pack measurements in April, 2008.

Proposed designs and applications

Heavy lifting

The proposed Aeroscraft is Aeros Corporation's continuation of the now canceled WALRUS project. This proposed craft is a hybrid airship that, while cruising, obtains two thirds of its lift from helium and the remaining third aerodynamic lift. Jets would be used during take-off and landing.

Passenger transport

There is a case for the airship or zeppelin as a medium to long distance air 'cruise ship' using helium as a lifting agent. Airship passengers could have spacious decks inside the hull to give ample room for sitting, sleeping and recreation. There would be ample room for restaurants and similar facilities. The potential exists for a market in more leisurely journeys, such as cruises over scenic terrain.

Practical comparison to fixed-wing aircraft

The advantage of airships over airplanes is that static lift sufficient for flight is generated by the lifting gas and requires no engine power. This was an immense advantage before the middle of WW I and remained an advantage for long distance, or long duration operations until WW II. Modern concepts for high altitude airships include photovoltaic cells to reduce the need to land to refuel, thus they can remain in the air until consumables expire.

The disadvantages are that the drag on an airship rises as the square of its speed, while the power required to propel it increases as the cube of the speed. In airplanes, lift and drag increase together with speed, so that for a given lift the drag is effectually constant at any speed, and so the power required only increases linearly with speed until close to the speed of sound. Given the large flat plate area and wetted surface of the airship, a practical limit is reached somewhere between 80 and 100 mph (160 km/h). So the airship is not trusted with an important position by speed, but by durability as surveillance-gathering platform or other airborne early warning mission. In these cases, speed in not a critical need.

The altitude an airship can fly at largely depends on how much lifting gas it can lose due to expansion before stasis is reached. The ultimate altitude record for a rigid airship was set in 1917 by the L-55 under the command of Kurt Flemming (who later died in the *Hindenburg*) when he forced the airship to 24,000 feet (7,300 m) attempting to cross France after the "Silent Raid" on London. The L-55 lost lift as the descent to lower altitudes over Germany compressed the gas left in the cells, and thus the weight of air displaced. L-55 crashed due to loss of lift. While such waste of gas was necessary for the survival of airships in the later years of WW I, it was impractical for commercial operations, or operations of helium filled military airships. The highest flight made by a hydrogen filled passenger airship was 5,500 feet on the *Graf Zeppelin's* around the world flight. The practical limits for rigid airships was about 3,000 feet (900 m), and for pressure airships around 8,000 feet (2,400 m).

Modern airships use dynamic helium volume. At sea level altitude, helium only takes up a small part of the hull, while the rest volume is filled with air. As the airship ascends, the helium inflates with reduced outer pressure, and air is pushed out and released from the downward valve. This allows airships to reach any altitude with balanced inner and outer pressure, if the buoyancy is enough. Some civil aerostat could reach 100912 feet without explosion due to overloaded inner pressure.

The greatest disadvantage of the airship is size, which is essential to increasing performance. As size increases, the problems of ground handling increase geometrically. As the German Navy transitioned from the "p" class Zeppelins of 1915 (with a volume of over 1.1 million cubic feet) to the larger "q" class of 1916, the "r" class of 1917, and finally the "w" class of 1918, at almost 2.2 million cubic feet, ground handling problems reduced the number of days the Zeppelins were able to make patrol flights. This availability declined from 34% in 1915, to 24.3% in 1916 and finally 17.5% in 1918.

So long as the power-to-weight ratios of aircraft engines remained low and specific fuel consumption high, the airship had an edge for long range or duration operations. As those figures changed, the balance shifted rapidly in the airplane's favour. By mid-1917 the airship could no longer survive in a combat situation where the threat was airplanes. By the late 1930s, the airship barely had an advantage over the airplane on intercontinental over-water flights, and that advantage had vanished by the end of WW II.

This is in face-to-face tactical situation, current High Altitude Airship project is planned to survey hundreds of kimometers as their operation radius, often much farther than normal engage range of a military airplane. This provides better early warning, even farther than the Aegis system. The current Aegis system is often based on sea vessel like Ticonderoga Class and Burke Class, which have restricted radio horizon and line of sight. For example, a radar mounted on a 30 meter high vessel platform has radio horizon at 19.5 kilometers range, while a radar mounted on an 18000m altitude HAA has radio horizon at 478.1 kilometers range. This is significantly important for detecting low-flying cruise missiles or fighter-bombers.

The blimp remained a viable military system only until the conventional submarine was replaced by the nuclear submarine. Today, airships are used primarily for command, control and communication platform; to establish and maintain reliable and secure connectivity among all forces, provide transparent data across the echelons; precisely locate friendly and enemy forces; detect targets on an extended battlefield at a minimal exposure to enemy forces; real time targeting; navigation assistance; battle management; monitor radio conversations, ect. These are not tough combat missions.

Safety

The lift gas, helium, is not merely inert but acts as a fire extinguisher.

Modern airships have a natural buoyancy and special design that offers a virtually zero catastrophic failure mode. The internal hull pressure is maintained at only 1%-2% above surrounding air pressure, the vehicle is highly tolerant to physical damage or to attack by small arms fire or missiles.

While on long-haul flights weather patterns would be flown to avoid bad weather, the hull's mass largely dampens the effect of turbulence – just as a large tanker rides through rough seas.

An airship is usually a poor lightning target, as it is constructed mainly from composite materials. If it is struck, built-in protection devices minimise the risk to the vehicle and its cargo.

A series of structural vulnerability tests were done by the UK Defence Evaluation and Research Agency DERA on a Skyship 600, an earlier airship built by the Munk team to a similar pressure-stabilised design. Several hundred high-velocity bullets were fired through the hull, and even two hours later the vehicle would have been able to return to base. The airship is virtually impervious to automatic rifle and mortar fire: ordnance passes through the envelope without causing critical helium loss. In all instances of light armament fire evaluated under both test and live conditions, the vehicle was able to complete its mission and return to base.

Alternating current

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

City lights viewed in a motion blurred exposure. The AC blinking causes the lines to be dotted rather than continuous.

An **alternating current** (**AC**) is an electrical current whose magnitude and direction vary cyclically, as opposed to direct current, whose direction remains constant. The usual waveform of an AC power circuit is a sine wave, as this results in the most efficient transmission of energy. However in certain applications different waveforms are used, such as triangular or square waves.

Used generically, AC refers to the form in which electricity is delivered to businesses and residences. However, audio and radio signals carried on electrical wire are also examples of alternating current. In these applications, an important goal is often the recovery of information encoded (or modulated) onto the AC signal.

History

William Stanley, Jr. designed one of the first practical devices to transfer AC power efficiently between isolated circuits. Using pairs of coils wound on a common iron core, his design, called an induction coil, was an early transformer. The AC power system system used today developed rapidly after 1886, and includes key concepts by Nikola Tesla, who subsequently sold his patent to George Westinghouse. Lucien Gaulard, John Dixon Gibbs, Carl Wilhelm Siemens and others contributed subsequently to this field. AC systems overcame the limitations of the direct current system used by Thomas Edison to distribute electricity efficiently over long distances.

The first modern commercial power plant using three-phase alternating current was at the Mill Creek hydroelectric plant near Redlands, California in 1893 designed by Almirian Decker. Decker's design incorporated 10,000 volt three-phase transmission and established the standards for the complete system of generation, transmission and motors used today.

Alternating current circuit theory evolved rapidly in the latter part of the 19th and early 20th century. Notable contributors to the theoretical basis of alternating current calculations include Charles Steinmetz, James Clerk Maxwell, Oliver Heaviside, and many others. Calculations in unbalanced three-phase systems were simplified by the symmetrical components methods discussed by Charles Legeyt Fortescue in 1918.

Transmission, distribution, and domestic power supply

AC power can be increased or decreased in voltage with a transformer. Use of a higher voltage leads to significantly more efficient transmission of power. The power losses in a conductor are a product of the square of the current and the resistance of the conductor, described by the formula $P = I^2 \cdot R$. This means that when transmitting a fixed power on a given wire, if the current is doubled, the power loss will be four times greater.

Since the power transmitted is equal to the product of the current, the voltage and the cosine of the phase difference φ ($P = IV\cos\varphi$), the same amount of power can be transmitted with a lower current by increasing the voltage. Therefore it is advantageous when transmitting large amounts of power to distribute the power with high voltages (often hundreds of kilovolts).

However, high voltages also have disadvantages, the main ones being the increased insulation required, and generally increased difficulty in their safe handling. In a power plant, power is generated at a convenient voltage for the design of a generator, and then stepped up to a high voltage for transmission. Near the loads, the transmission voltage is stepped down to the voltages used by equipment. Consumer voltages vary depending on the country and size of load, but generally motors and lighting are built to use up to a few hundred volts between phases.

The utilization voltage delivered to equipment such as lighting and motor loads is standardized, with an allowable range of voltage over which equipment is expected to operate. Standard power utilization voltages and percentage tolerance vary in the different mains power systems found in the world.

Modern high-voltage, direct-current electric power transmission systems contrast with the more common alternating-current systems as a means for the bulk transmission of electrical power over long distances. HVDC systems tend to be more expensive and less efficient than transformers. Transmission with high voltage direct current was not feasible when Edison, Westinghouse and Tesla were designing their power systems, since there was then no way to economically convert AC power to DC and back again at the necessary voltages.

Three-phase electrical generation is very common. Three separate coils in the generator stator are physically offset by an angle of 120° to each other. Three current waveforms are produced that are equal in magnitude and 120° out of phase to each other.

If the load on a three-phase system is balanced equally among the phases, no current flows through the neutral point. Even in the worst-case unbalanced (linear) load, the neutral current will not exceed the highest of the phase currents. It is noteworthy that non-linear loads (e.g. computers) may require an oversized neutral bus and neutral conductor in the upstream distribution panel to handle harmonics. Harmonics can cause neutral conductor current levels to exceed that of one or all phase conductors.

For three-phase at utilization voltages a four-wire system is often used. When stepping down three-phase, a transformer with a Delta primary and a Star secondary is often used so there is no need for a neutral on the supply side.

For smaller customers (just how small varies by country and age of the installation) only a single phase and the neutral or two phases and the neutral are taken to the property. For larger installations all three phases and the neutral are taken to the main distribution panel. From the three-phase main panel, both single and three-phase circuits may lead off.

Three-wire single phase systems, with a single centre-tapped transformer giving two live conductors, is a common distribution scheme for residential and small commercial buildings in North America. This arrangement is sometimes incorrectly referred to as "two phase". A similar method is used for a different reason on construction sites in the UK. Small power tools and lighting are supposed to be supplied by a local centre-tapped transformer with a voltage of 55V between each power conductor and the earth. This significantly reduces the risk of electric shock in the event that one of the live conductors becomes exposed through an equipment fault whilst still allowing a reasonable voltage for running the tools.

A third wire, called the bond wire, is often connected between non-current carrying metal enclosures and earth ground. This conductor provides protection from electrical shock due to accidental contact of circuit conductors with the metal chassis of portable appliances and tools. Bonding all non-current carrying metal parts into one complete system ensures there is always a low impedance path to ground sufficient to carry any fault current for as long as it takes for the system to clear the fault. This low impedance path allows the maximum amount of fault current to flow, causing the overcurrent protection device (Breakers, fuses) to trip or burn out as quickly as possible, returning the electrical system to a safe state. All bond wires are bonded to ground at the main service panel, as is the Neutral/Identified Conductor if present.

AC power supply frequencies

The frequency of the electrical system varies by country; most electric power is generated at either 50 or 60 Hz. See List of countries with mains power plugs, voltages and frequencies. Some countries have a mixture of 50 Hz and 60 Hz supplies, notably Japan.

A low frequency eases the design of low speed electric motors, particularly for hoisting, crushing and rolling applications, and commutator-type traction motors for applications such as railways, but also causes a noticeable flicker in incandescent lighting and objectionable flicker of fluorescent lamps. 16⅔ Hz power is still used in some European rail systems, such as in Austria, Germany, Norway, Sweden and Switzerland. The use of lower frequencies also provided the advantage of lower impedance losses, which are proportional to frequency. The original Niagara Falls generators were built to produce 25 Hz power, as a compromise between low frequency for traction and heavy induction motors, while still allowing incandescent lighting to operate (although with noticeable flicker); most of the 25 Hz residential and commercial customers for Niagara Falls power were converted to 60 Hz by the late 1950's, although some 25 Hz industrial customers still existed as of the start of the 21st century.

Off-shore,military, textile industry, marine, computer mainframe, aircraft, and spacecraft applications sometimes use 400 Hz, for benefits of reduced weight of apparatus or higher motor speeds.

Effects at high frequencies

A direct, constant current flows uniformly throughout the cross-section of the (uniform) wire that carries it. With alternating current of any frequency, the current is forced towards the outer surface of the wire, and away from the centre. This is because an electric charge which accelerates (as is the case of an alternating current) radiates electromagnetic waves, and materials of high conductivity (the metal which makes up the wire) do not allow propagation of electromagnetic waves. This phenomenon is called skin effect.

At very high frequencies the current no longer flows *in* the wire, but effectively flows *on* the surface of the wire, within a thickness of a few skin depths. The skin depth is the thickness at which the current density is reduced by 63%. Even at relatively low frequencies used for high power transmission (50–60 Hz), non-uniform distribution of current still occurs in sufficiently thick conductors. For example, the skin depth of a copper conductor is approximately 8.57 mm at 60 Hz, so high current conductors are usually hollow to reduce their mass and cost.

Since the current tends to flow in the periphery of conductors, the effective cross-section of the conductor is reduced. This increases the effective *AC* resistance of the conductor, since resistance is inversely proportional to the cross-sectional area in which the current actually flows. The AC resistance often is many times higher than the *DC* resistance, causing a much higher energy loss due to ohmic heating (also called I^2R loss).

Techniques for reducing AC resistance

For low to medium frequencies, conductors can be divided into stranded wires, each insulated from one other, and the individual strands specially arranged to change their relative position within the conductor bundle. Wire constructed using this technique is called Litz wire. This measure helps to partially mitigate skin effect by forcing more equal current flow throughout the total cross section of the stranded conductors. Litz wire is used for making high Q inductors, reducing losses in flexible conductors carrying very high currents at power frequencies, and in the windings of devices carrying higher radio frequency current (up to hundreds of kilohertz), such as switch-mode power supplies and radio frequency transformers.

Techniques for reducing radiation loss

As written above, an alternating current is made of electric charge under periodic acceleration, which causes radiation of electromagnetic waves. Energy that is radiated represents a loss. Depending on the frequency, different techniques are used to minimize the loss due to radiation.

Twisted pairs

At frequencies up to about 1 GHz, wires are paired together in cabling to form a twisted pair in order to reduce losses due to electromagnetic radiation and inductive coupling. A twisted pair must be used with a balanced signalling system, where the two wires carry equal but opposite currents. The result is that each wire in the twisted pair radiates a signal that is effectively cancelled by the other wire, resulting in almost no electromagnetic radiation.

Coaxial cables

At frequencies above 1 GHz, unshielded wires of practical dimensions lose too much energy to radiation, so coaxial cables are used instead. A coaxial cable has a conductive wire inside a conductive tube. The current flowing on the inner conductor is equal and opposite to the current flowing on the inner surface of the outer tube. This causes the electromagnetic field to be completely contained within the tube, and (ideally) no energy is radiated or coupled outside the tube. Coaxial cables have acceptably small losses for frequencies up to about 20 GHz. For microwave frequencies greater than 20 GHz, the dielectric losses (due mainly to the dissipation factor of the dielectric layer which separates the inner wire from the outer tube) become too large, making waveguides a more efficient medium for transmitting energy.

Waveguides

Waveguides are similar to coax cables, as both consist of tubes, with the biggest difference being that the waveguide has no inner conductor. Waveguides can have any arbitrary cross section, but rectangular cross sections are the most common. With waveguides, the energy is no longer carried by an electric current, but by a *guided* electromagnetic field. Waveguides have dimensions comparable to the wavelength of the alternating current to be transmitted, so they are only feasible at microwave frequencies.

Fibre optics

At frequencies greater than 200 GHz, waveguide dimensions become impractically small, and the ohmic losses in the waveguide walls become large. Instead, fibre optics, which are a form of dielectric waveguides, can be used. For such frequencies, the concepts of voltages and currents are no longer used.

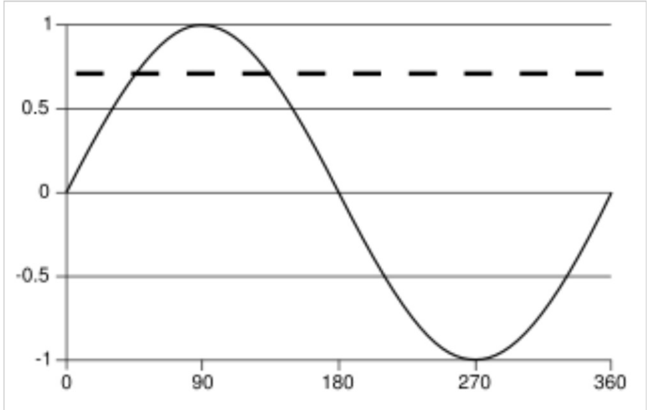
Mathematics of AC voltages

Alternating currents are accompanied (or caused) by alternating voltages. In English the initialism AC is commonly and somewhat confusingly used for both. An AC voltage *v* can be described mathematically as a function of time by the following equation:

$$v(t) = V_{\mathrm{peak}} \cdot \sin(\omega t),$$

where

- V_{peak} is the peak voltage (unit: volt),
- ω is the angular frequency (unit: radians per second)
  - The angular frequency is related to the physical frequency, f, which represents the number of oscillations per second (unit = hertz), by the equation $\omega = 2\pi f$.
- t is the time (unit: second).



A sine wave, over one cycle (360°). The dashed line represents the root mean square (RMS) value at about 0.707

The peak-to-peak value of an AC voltage is defined as the difference between its positive peak and its negative peak. Since the maximum value of $\sin(x)$ is +1 and the minimum value is −1, an AC voltage swings between $+V_{\mathrm{peak}}$ and $-V_{\mathrm{peak}}$. The peak-to-peak voltage, usually written as V_{pp} or $V_{\mathrm{P-P}}$, is therefore $V_{\mathrm{peak}} - (-V_{\mathrm{peak}}) = 2 \times V_{\mathrm{peak}}$.

Power and root mean square

The relationship between voltage and power is:

$$P(t) = \frac{V^2(t)}{R}$$ where R represents a load resistance

Rather than using instantaneous power, $P\left(t\right)$, it is more practical to use a time averaged power (where the averaging is performed over any integer number of cycles). Therefore, AC voltage is often expressed as a root mean square (RMS) value, written as $V_{\rm rms}$, because

$$P_{\rm time~averaged} = \frac{{V^2}_{\rm rms}}{R}$$

For a sinusoidal voltage:

$$V_{\rm rms} = \frac{V_{\rm peak}}{\sqrt{2}}$$

The factor $\sqrt{2}$ is called the crest factor, which varies for different waveforms.

- For a triangle wave form: $V_{\rm rms} = \frac{V_{\rm peak}}{\sqrt{3}}$
- For a square wave form: $V_{\rm rms} = V_{\rm peak}$

Example

To illustrate these concepts, consider a 240 V AC mains supply. It is so called because its Root mean square value is 240 V. This means that the time-averaged power delivered is equivalent to the power delivered by a DC voltage of 240 Volts. To determine the peak voltage (amplitude), we can modify the above equation to:

$$V_{\rm peak} = \sqrt{2}\ V_{\rm rms}$$

For our 240 V AC, the peak voltage $V_{\rm peak}$ is therefore $240V \times \sqrt{2}$, which is about 339 V. The peak-to-peak value V_{P-P} of the 240 V AC is double that, at about 679 V.

Retrieved from " http://en.wikipedia.org/wiki/Alternating_current"

Automobile

2008/9 Schools Wikipedia Selection. Related subjects: Road transport

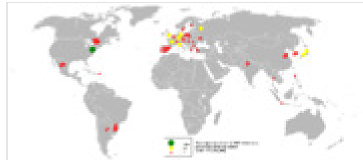
An **automobile** or **motor car** is a wheeled motor vehicle for transporting passengers; which also carries its own engine or motor. Most definitions of the term specify that automobiles are designed to run primarily on roads, to have seating for one to eight people, to typically have four wheels, and to be constructed principally for the transport of people rather than goods. However, the term is far from precise because there are many types of vehicles that do similar tasks.


Karl Benz's "Velo" model (1894) - entered into an early automobile race

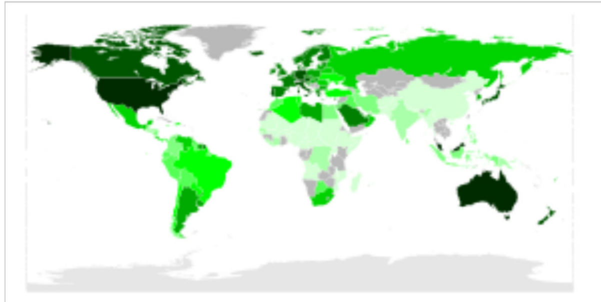
Automobile comes via the French language, from the Greek language by combining *auto* [self] with *mobilis* [moving]; meaning a vehicle that moves itself, rather than being pulled or pushed by a separate animal or another vehicle. The alternative name *car* is believed to originate from the Latin word *carrus* or *carrum* [wheeled vehicle], or the Middle English word *carre* [cart] (from Old North French), and *karros*; a Gallic wagon.

As of 2002, there were 590 million passenger cars worldwide (roughly one car per eleven people).

History


Passenger cars in 2000

Although Nicolas-Joseph Cugnot is often credited with building the first self-propelled mechanical vehicle or automobile in about 1769 by adapting an existing horse-drawn vehicle, this claim is disputed by some, who doubt Cugnot's three-wheeler ever ran or was stable. Others claim Ferdinand Verbiest, a member of a Jesuit mission in China, built the first steam-powered vehicle around 1672 which was of small scale and designed as a toy for the Chinese Emperor that was unable to carry a driver or a passenger, but quite possibly, was the first working steam-powered vehicle ('auto-mobile'). What is not in doubt is that Richard Trevithick built and demonstrated his *Puffing Devil* road locomotive in 1801, believed by many to be the first demonstration of a steam-powered road vehicle although it was unable to maintain sufficient steam pressure for long periods, and would have been of little practical use.


World map of passenger cars per 1000 people.

In Russia in the 1780s Ivan Kulibin started working on a human-pedalled carriage with s steam engine. He finished working on it in 1791. Some of its features included a flywheel, brake, gear box, and bearing, which are also the features of a modern automobile. His design had three wheels. Unfortunately, like for many of his inventions, the government failed to see the potential market and it was not developed further.

François Isaac de Rivaz, a Swiss inventor, designed the first internal combustion engine, in 1806, which was fueled by a mixture of hydrogen and oxygen and used it to develop the world's first vehicle, albeit rudimentary, to be powered by such an engine. The design was not very successful, as was the case with those of Samuel Brown, Samuel Morey, and Etienne Lenoir who each produced vehicles (adapted carriages, carts, or boats) powered by clumsy internal combustion engines.

In November 1881 French inventor Gustave Trouvé demonstrated a working three-wheeled automobile that was powered by electricity. This was at the International Exhibition of Electricity in Paris.

Although several other German engineers (including Gottlieb Daimler, Wilhelm Maybach, and Siegfried Marcus) were working on the problem at about the same time, **Karl Benz** generally is acknowledged as the inventor of the modern automobile.

An automobile powered by his own four-stroke cycle gasoline engine was built in Mannheim, Germany by Karl Benz in 1885 and granted a patent in January of the following year under the auspices of his major company, Benz & Cie., which was founded in 1883. It was an integral design, without the adaptation of other existing components and including several new technological elements to create a new concept. This is what made it worthy of a patent. He began to sell his production vehicles in 1888.

Karl Benz

In 1879 Benz was granted a patent for his first engine, which had been designed in 1878. Many of his other inventions made the use of the internal combustion engine feasible for powering a vehicle.

A photograph of the original *Benz Patent Motorwagon*, first built in 1885 and awarded the patent for the concept

His first *Motorwagon* was built in 1885 and he was awarded the patent for its invention as of his application on January 29, 1886. Benz began promotion of the vehicle on July 3, 1886 and approximately 25 Benz vehicles were sold between 1888 and 1893, when his first four-wheeler was introduced along with a model intended for affordability. They also were powered with four-stroke engines of his own design. Emile Roger of France, already producing Benz engines under license, now added the Benz automobile to his line of products. Because France was more open to the early automobiles, initially more were built and sold in France through Roger than Benz sold in Germany.

In 1896, Benz designed and patented the first internal-combustion flat engine, called a *boxermotor* in German. During the last years of the nineteenth century, Benz was the largest automobile company in the world with 572 units produced in 1899 and because of its size, Benz & Cie., became a joint-stock company.

Daimler and Maybach founded Daimler Motoren Gesellschaft (Daimler Motor Company, DMG) in Cannstatt in 1890 and under the brand name, *Daimler*, sold their first automobile in 1892, which was a horse-drawn stagecoach built by another manufacturer, that they retrofitted with an engine of their design. By 1895 about 30 vehicles had been built by Daimler and Maybach, either at the Daimler works or in the Hotel Hermann, where they set up shop after falling out with their backers. Benz and the Maybach and Daimler team seem to have been unaware of each other's early work. They never worked together because by the time of the merger of the two companies, Daimler and Maybach were no longer part of DMG.

Daimler died in 1900 and later that year, Maybach designed an engine named *Daimler-Mercedes*, that was placed in a specially-ordered model built to specifications set by Emil Jellinek. This was a production of a small number of vehicles for Jellinek to race and market in his country. Two years later, in 1902, a new model DMG automobile was produced and the model was named Mercedes after the Maybach engine which generated 35 hp. Maybach quit DMG shortly thereafter and opened a business of his own. Rights to the *Daimler* brand name were sold to other manufacturers.

Karl Benz proposed co-operation between DMG and Benz & Cie. when economic conditions began to deteriorate in Germany following the First World War, but the directors of DMG refused to consider it initially. Negotiations between the two companies resumed several years later when these conditions worsened and, in 1924 they signed an *Agreement of Mutual Interest*, valid until the year 2000. Both enterprises standardized design, production, purchasing, and sales and they advertised or marketed their automobile models jointly—although keeping their respective brands.

On June 28, 1926, Benz & Cie. and DMG finally merged as the *Daimler-Benz* company, baptizing all of its automobiles *Mercedes Benz* as a brand honoring the most important model of the DMG automobiles, the Maybach design later referred to as the *1902 Mercedes-35hp*, along with the Benz name. Karl Benz remained a member of the board of directors of Daimler-Benz until his death in 1929 and at times, his two sons participated in the management of the company as well.

In 1890, Emile Levassor and Armand Peugeot of France began producing vehicles with Daimler engines and so laid the foundation of the automobile industry in France.

The first design for an American automobile with a gasoline internal combustion engine was drawn in 1877 by George Selden of Rochester, New York, who applied for a patent for an automobile in 1879, but the patent application expired because the vehicle was never built and proved to work (a requirement for a patent). After a delay of sixteen years and a series of attachments to his application, on November 5, 1895, Selden was granted a United States patent () for a two-stroke automobile engine, which hindered, more than encouraged, development of automobiles in the United States. His patent was challenged by Henry Ford and others, and overturned in 1911.

In Britain there had been several attempts to build steam cars with varying degrees of success with Thomas Rickett even attempting a production run in 1860. Santler from Malvern is recognized by the Veteran Car Club of Great Britain as having made the first petrol-powered car in the country in 1894 followed by Frederick William Lanchester in 1895 but these were both one-offs. The first production vehicles in Great Britain came from the Daimler Motor Company, a company founded by Harry J. Lawson in 1896 after purchasing the right to use the name of the engines. Lawson's company made its first automobiles in 1897 and they bore the name *Daimler*.

In 1892, German engineer Rudolf Diesel was granted a patent for a "New Rational Combustion Engine". In 1897 he built the first Diesel Engine. Steam-, electric-, and gasoline-powered vehicles competed for decades, with gasoline internal combustion engines achieving dominance in the 1910s.

Although various pistonless rotary engine designs have attempted to compete with the conventional piston and crankshaft design, only Mazda's version of the Wankel engine has had more than very limited success.

Production

The large-scale, production-line manufacturing of affordable automobiles was debuted by Ransom Olds at his Oldsmobile factory in 1902. This concept was greatly expanded by Henry Ford, beginning in 1914.


Ransom E. Olds.

As a result, Ford's cars came off the line in fifteen minute intervals, much faster than previous methods, increasing production by seven to one (requiring 12.5 man-hours before, 1 hour 33 minutes after), while using less manpower. It was so successful, paint became a bottleneck. Only Japan black would dry fast enough, forcing the company to drop the variety of colors available before 1914, until fast-drying Duco lacquer was developed in 1926. This is the source of Ford's apocryphal remark, "any colour as long as it's black". In 1914, an assembly line worker could buy a Model T with four months' pay.


Portrait of Henry Ford (ca. 1919)

Ford's complex safety procedures—especially assigning each worker to a specific location instead of allowing them to roam about—dramatically reduced the rate of injury. The combination of high wages and high efficiency is called " Fordism," and was copied by most major industries. The efficiency gains from the assembly line also coincided with the economic rise of the United States. The assembly line forced workers to work at a certain pace with very repetitive motions which led to more output per worker while other countries were using less productive methods.

In the automotive industry, its success was dominating, and quickly spread worldwide seeing the founding of Ford France and Ford Britain in 1911, Ford Denmark 1923, Ford Germany 1925; in 1921, Citroen was the first native European manufacturer to adopt the production method. Soon, companies had to have assembly lines, or risk going broke; by 1930, 250 companies which did not, had disappeared.

Development of automotive technology was rapid, due in part to the hundreds of small manufacturers competing to gain the world's attention. Key developments included electric ignition and the electric self-starter (both by Charles Kettering, for the Cadillac Motor Company in 1910-1911), independent suspension, and four-wheel brakes.

Since the 1920s, nearly all cars have been mass-produced to meet market needs, so marketing plans often have heavily influenced automobile design. It was Alfred P. Sloan who established the idea of different makes of cars produced by one company, so buyers could "move up" as their fortunes improved.

Ford Model T, 1927, regarded as the first affordable American automobile

Reflecting the rapid pace of change, makes shared parts with one another so larger production volume resulted in lower costs for each price range. For example, in the 1930s, LaSalles, sold by Cadillac, used cheaper mechanical parts made by Oldsmobile; in the 1950s, Chevrolet shared hood, doors, roof, and windows with Pontiac; by the 1990s, corporate drivetrains and shared platforms (with interchangeable brakes, suspension, and other parts) were common. Even so, only major makers could afford high costs, and even companies with decades of production, such as Apperson, Cole, Dorris, Haynes, or Premier, could not manage: of some two hundred American car makers in existence in 1920, only 43 survived in 1930, and with the Great Depression, by 1940, only 17 of those were left.

In Europe much the same would happen. Morris set up its production line at Cowley in 1924, and soon outsold Ford, while beginning in 1923 to follow Ford's practise of vertical integration, buying Hotchkiss (engines), Wrigley (gearboxes), and Osberton (radiators), for instance, as well as competitors, such as Wolseley: in 1925, Morris had 41% of total British car production. Most British small-car assemblers, from Abbey to Xtra had gone under. Citroen did the same in France, coming to cars in 1919; between them and other cheap cars in reply such as Renault's 10CV and Peugeot's 5CV, they produced 550,000 cars in 1925, and Mors, Hurtu, and others could not compete. Germany's first mass-manufactured car, the Opel 4PS *Laubfrosch* (Tree Frog), came off the line at Russelsheim in 1924, soon making Opel the top car builder in Germany, with 37.5% of the market.

Fuel and propulsion technologies

Most automobiles in use today are propelled by gasoline (also known as petrol) or diesel internal combustion engines, which are known to cause air pollution and are also blamed for contributing to climate change and global warming. Increasing costs of oil-based fuels, tightening environmental laws and restrictions on greenhouse gas emissions are propelling work on alternative power systems for automobiles. Efforts to improve or replace existing technologies include the development of hybrid vehicles, and electric and hydrogen vehicles which do not release pollution into the air.

Auto rickshaws in New Delhi run on Compressed Natural Gas

Diesel

Diesel-engined cars have long been popular in Europe with the first models being introduced in the 1930s by Mercedes Benz and Citroen. The main benefit of diesel engines is a 50% fuel burn efficiency compared with 27% in the best gasoline engines. A down-side of the diesel is the presence in the exhaust gases of fine soot particulates and manufacturers are now starting to fit filters to remove these. Many diesel-powered cars can also run with little or no modifications on 100% biodiesel.

A CNG powered high-floor Neoplan AN440A, run on Compressed Natural Gas

Gasoline

Gasoline engines have the advantage over diesel in being lighter and able to work at higher rotational speeds and they are the usual choice for fitting in high-performance sports cars. Continuous development of gasoline engines for over a hundred years has produced improvements in efficiency and reduced pollution. The carburetor was used on nearly all road car engines until the 1980s but it was long realised better control of the fuel/air mixture could be achieved with fuel injection. Indirect fuel injection was first used in aircraft engines from 1909, in racing car engines from the 1930s, and road cars from the late 1950s. Gasoline Direct Injection (GDI) is now starting to appear in production vehicles such as the 2007 (Mark II) BMW Mini. Exhaust gases are also cleaned up by fitting a catalytic converter into the exhaust system. Clean air legislation in many of the car industries most important markets has made both catalysts and fuel injection virtually universal fittings. Most modern gasoline engines also are capable of running with up to 15% ethanol mixed into the gasoline - older vehicles may have seals and hoses that can be harmed by ethanol. With a small amount of redesign, gasoline-powered vehicles can run on ethanol concentrations as high as 85%. 100% ethanol is used in some parts of the world (such as Brazil), but vehicles must be started on pure gasoline and switched over to ethanol once the engine is running. Most gasoline engined cars can also run on LPG with the addition of an LPG tank for fuel storage and carburetion modifications to add an LPG mixer. LPG produces fewer toxic emissions and is a popular fuel for fork lift trucks that have to operate inside buildings.

2007 Mark II (BMW) Mini Cooper

Bioalcohols and biogasoline

Ethanol, other alcohol fuels (biobutanol) and biogasoline have widespread use an automotive fuel. Most alcohols have less energy per liter than gasoline and are usually blended with gasoline. Alcohols are used for a variety of reasons - to increase octane, to improve emissions, and as an alternative to petroleum based fuel, since they can be made from agricultural crops. Brazil's ethanol program provides about 20% of the nations automotive fuel needs, including several million cars that operate on pure ethanol.

The hydrogen powered FCHV (Fuel Cell Hybrid Vehicle) was developed by Toyota in 2005

Electric

The first electric cars were built around 1832, well before internal combustion powered cars appeared. For a period of time electrics were considered superior due to the silent nature of electric motors compared to the very loud noise of the gasoline engine. This advantage was removed with Hiram Percy Maxim's invention of the muffler in 1897. Thereafter internal combustion powered cars had two critical advantages: 1) long range and 2) high specific energy (far lower weight of petrol fuel versus weight of batteries). The building of battery electric vehicles that could rival internal combustion models had to wait for the introduction of modern semiconductor controls and improved batteries. Because they can deliver a high torque at low revolutions electric cars do not require such a complex drive train and transmission as internal combustion powered cars. Some post-2000 electric car designs such as the Venturi Fétish are able to accelerate from 0-60 mph (96 km/h) in 4.0 seconds with a top speed around 130 mph (210 km/h). Others have a range of 250 miles (400 km) on the EPA highway cycle requiring 3-1/2 hours to completely charge. Equivalent fuel efficiency to internal combustion is not well defined but some press reports give it at around 135 mpg–U.S. (1.74 L/100 km / 162.1 mpg–imp).

The Henney Kilowatt, the first modern (transistor-controlled) electric car.

Steam

Steam power, usually using an oil- or gas-heated boiler, was also in use until the 1930s but had the major disadvantage of being unable to power the car until boiler pressure was available (although the newer models could achieve this in well under a minute). It has the advantage of being able to produce very low emissions as the combustion process can be carefully controlled. Its disadvantages include poor heat efficiency and extensive requirements for electric auxiliaries.

2007 Tesla electric powered Roadster

Air

A compressed air car is an alternative fuel car that uses a motor powered by compressed air. The car can be powered solely by air, or by air combined (as in a hybrid electric vehicle) with gasoline/diesel/ethanol or electric plant and regenerative braking. Instead of mixing fuel with air and burning it to drive pistons with hot expanding gases; *compressed air cars* use the expansion of compressed air to drive their pistons. Several prototypes are available already and scheduled for worldwide sale by the end of 2008. Companies releasing this type of car include Tata Motors and Motor Development International (MDI).


Tata/MDI OneCAT Air Car

Gas turbine

In the 1950s there was a brief interest in using gas turbine (jet) engines and several makers including Rover and Chrysler produced prototypes. In spite of the power units being very compact, high fuel consumption, severe delay in throttle response, and lack of engine braking meant no cars reached production.

Rotary (Wankel) engines

Rotary Wankel engines were introduced into road cars by NSU with the Ro 80 and later were seen in the Citroën GS Birotor and several Mazda models. In spite of their impressive smoothness, poor reliability and fuel economy led to them largely disappearing. Mazda, beginning with the R100 then RX-2, has continued research on these engines, overcoming most of the earlier problems with the RX-7 and RX-8.

Rocket and jet cars

A rocket car holds the record in drag racing. However, the fastest of those cars are used to set the Land Speed Record, and are propelled by propulsive jets emitted from rocket, turbojet, or more recently and most successfully turbofan engines. The ThrustSSC car using two Rolls-Royce Spey turbofans with reheat was able to exceed the speed of sound at ground level in 1997.

Safety


Result of a serious automobile accident.

Road traffic injuries represent about 25% of worldwide injury-related deaths (the leading cause) with an estimated 1.2 million deaths (2004) each year.

Automobile accidents are almost as old as automobiles themselves. Early examples include Mary Ward, who became one of the first documented automobile fatalities in 1869 in Parsonstown, Ireland, and Henry Bliss, one of the United State's first pedestrian automobile casualties in 1899 in New York.

Cars have many basic safety problems - for example, they have human drivers who can make mistakes, wheels that can lose traction when braking, turning or acceleration forces are too high, and mechanical systems subject to failure. Collisions can have very serious or fatal consequences. Some vehicles have a high centre of gravity and therefore an increased tendency to roll over.

Early safety research focused on increasing the reliability of brakes and reducing the flammability of fuel systems. For example, modern engine compartments are open at the bottom so that fuel vapors, which are heavier than air, vent to the open air. Brakes are hydraulic and dual circuit so that a total braking failure is very rare. Systematic research on crash safety started in 1958 at Ford Motor Company. Since then, most research has focused on absorbing external crash energy with crushable panels and reducing the motion of human bodies in the passenger compartment. This is reflected in most cars produced today.

Significant reductions in death and injury have come from the addition of Safety belts and laws in many countries to require vehicle occupants to wear them. Airbags and specialised child restraint systems have improved on that. Structural changes such as side-impact protection bars in the doors and side panels of the car mitigate the effect of impacts to the side of the vehicle. Many cars now include radar or sonar detectors mounted to the rear of the car to warn the driver if he or she is about to reverse into an obstacle or a pedestrian. Some vehicle manufacturers are producing cars with devices that also measure the proximity to obstacles and other vehicles in front of the car and are using these to apply the brakes when a collision is inevitable. There have also been limited efforts to use heads up displays and thermal imaging technologies similar to those used in military aircraft to provide the driver with a better view of the road at night.

There are standard tests for safety in new automobiles, like the EuroNCAP and the US NCAP tests. There are also tests run by organizations such as IIHS and backed by the insurance industry.

Despite technological advances, there is still significant loss of life from car accidents: About 40,000 people die every year in the United States, with similar figures in European nations. This figure increases annually in step with rising population and increasing travel if no measures are taken, but the rate *per capita* and *per* mile traveled decreases steadily. The death toll is expected to nearly double worldwide by 2020. A much higher number of accidents result in injury or permanent disability. The highest accident figures are reported in China and India. The European Union has a rigid program to cut the death toll in half by 2010, and member states have started implementing measures.

Automated control has been seriously proposed and successfully prototyped. Shoulder-belted passengers could tolerate a 32 *g* emergency stop (reducing the safe inter-vehicle gap 64-fold) if high-speed roads incorporated a steel rail for emergency braking. Both safety modifications of the roadway are thought to be too expensive by most funding authorities, although these modifications could dramatically increase the number of vehicles able to safely use a high-speed highway. This makes clear the often-ignored fact road design and traffic control also play a part in car wrecks; unclear traffic signs, inadequate signal light placing, and poor planning (curved bridge approaches which become icy in winter, for example), also contribute.

Economics and impacts

Cost and benefits of usage

The costs of automobile usage, which may include the cost of: acquiring the vehicle, repairs, maintenance, fuel, depreciation, parking fees, tire replacement, taxes and insurance, are weighed against the cost of the alternatives, and the value of the benefits - perceived and real - of vehicle usage. The benefits may include on-demand transportation, mobility, independence and convenience.

Cost and benefits to society

Similarly the costs to society of encompassing automobile use, which may include those of: maintaining roads, land use, pollution, public health, health care, and of disposing of the vehicle at the end of its life, can be balanced against the value of the benefits to society that automobile use generates. The societal benefits may include: economy benefits, such as job and wealth creation, of automobile production and maintenance, transportation provision, society wellbeing derived from leisure and travel opportunities, and revenue generation from the tax opportunities. The ability for humans to move flexibly from place to place has far reaching implications for the nature of societies.

Impacts on society and environment

Transportation is a major contributor to air pollution in most industrialised nations. According to the American Surface Transportation Policy Project nearly half of all Americans are breathing unhealthy air. Their study showed air quality in dozens of metropolitan areas has got worse over the last decade. In the United States the average passenger car emits 11,450 lbs (5 tonnes) of carbon dioxide, along with smaller amounts of carbon monoxide, hydrocarbons, and nitrogen. Residents of low-density, residential-only sprawling communities are also more likely to die in car collisions, which kill 1.2 million people worldwide each year, and injure about forty times this number. Sprawl is more broadly a factor in inactivity and obesity, which in turn can lead to increased risk of a variety of diseases.

Improving the positive and reducing the negative impacts

Fuel taxes may act as an incentive for the production of more efficient, hence less polluting, car designs (e.g. hybrid vehicles) and the development of alternative fuels. High fuel taxes may provide a strong incentive for consumers to purchase lighter, smaller, more fuel-efficient cars, or to not drive. On average, today's automobiles are about 75 percent recyclable, and using recycled steel helps reduce energy use and pollution. In the United States Congress, federally mandated fuel efficiency standards have been debated regularly, passenger car standards have not risen above the 27.5 mpg–U.S. (8.55 L/100 km / 33 mpg–imp) standard set in 1985. Light truck standards have changed more frequently, and were set at 22.2 mpg–U.S. (10.6 L/100 km / 26.7 mpg–imp) in 2007. Alternative fuel vehicles are another option that is less polluting than conventional petroleum powered vehicles.

Future car technologies

Automobile propulsion technology under development include gasoline/electric and plug-in hybrids, battery electric vehicles, hydrogen cars, biofuels, and various alternative fuels.

Research into future alternative forms of power include the development of fuel cells, Homogeneous Charge Compression Ignition (HCCI), stirling engines, and even using the stored energy of compressed air or liquid nitrogen.

New materials which may replace steel car bodies include duraluminum, fibreglass, carbon fibre, and carbon nanotubes.

Telematics technology is allowing more and more people to share cars, on a pay-as-you-go basis, through such schemes as City Car Club in the UK, Mobility in mainland Europe, and Zipcar in the US.

Alternatives to the automobile

Established alternatives for some aspects of automobile use include public transit (buses, trolleybuses, trains, subways, monorails, tramways), cycling, walking, rollerblading, skateboarding and using a velomobile. Car-share arrangements and carpooling are also increasingly popular–the U.S. market leader in car-sharing has experienced double-digit growth in revenue and membership growth between 2006 and 2007, offering a service that enables urban residents to "share" a vehicle rather than own a car in already congested neighborhoods. Bike-share systems have been tried in some European cities, including Copenhagen and Amsterdam. Similar programs have been experimented with in a number of U.S. Cities. Additional individual modes of transport, such as personal rapid transit could serve as an alternative to automobiles if they prove to be socially accepted.

Axe

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Axe

The **axe**, or **ax**, is an implement that has been used for millennia to shape, split and cut wood, harvest timber, as a weapon and a ceremonial or heraldic symbol. The axe has many forms and specialized uses but generally consists of an axe head with a handle, or *helve*.

The earliest examples of axes have heads of stone with some form of wooden handle attached (hafted) in a method to suit the available materials and use. Axes made of copper, bronze, iron, steel appeared as these technologies developed. The axe is an example of a simple machine, as it is a type of wedge, or dual inclined plane. This reduces the effort needed by the wood chopper. It splits the wood into two parts by the pressure concentration at the blade. The handle of the axe also acts as a lever allowing the user to increase the force at the cutting edge (try using an axe head without a handle and you will see what is meant) - not using the full length of the handle is known as choking the axe. For fine chopping using a side axe this sometimes is a positive effect, but for felling with a double bitted axe it reduces efficiency.

axe of Stihl Holding AG & Co. KG

Generally cutting axes have a shallow wedge angle, whereas splitting axes have a deeper angle. Most axes are double beveled, i.e. symmetrical about the axis of the blade, but some specialist broadaxes have a single bevel blade, and usually an offset handle that allows them to be used for finishing work without putting the user's knuckles at risk of injury. Less common today they were once an integral part of a joiner and carpenter's tool kit - not just a tool for use in forestry. A tool of similar origin is the billhook with short handle and long blade it can be used for tasks where an axe is unsuitable. However in France and Holland the billhook often replaced the axe as a joiner's bench tool.

Most modern axes have steel heads and wooden handles, typically hickory in the USA and ash in the UK and Europe, although plastic or fibreglass handles are also common. Modern axes are specialized by use, size and form. Hafted axes with short handles designed for use with one hand are often called hand axes but the term hand axe refers to axes without handles as well. Hatchets tend to be small hafted axes often with a hammer on the back side (the poll).

Axes were frequently used in combat as they were easy to make, and the village edge tool makers were frequently the armourers to the lord of the manor in times of war.

History

Initially axes were probably not hafted. The first true hafted axes are known from the Mesolithic period (ca. 6000 BC). Axes made from ground stone are known since the Neolithic. Few wooden hafts have been found from this period, but it seems that the axe was normally hafted by wedging. Birch-tar and raw-hide lashings were used to fix the blade.

Roman axe in an ancient Roman relief in Brescia, Italy

Sometimes a short section of deer antler (an "antler sleeve") was used, which prevented the splitting of the haft and softened the impact on the stone blade itself, helping absorb the impact of each axe blow and lessening the chances of breaking the handle. The antler was hollowed out at one end to create a socket for the axehead. The antler sheath was then either perforated and a handle inserted into it or set in a hole made in the handle instead.

The distribution of stone axes is an important indication of prehistoric trade. thin sectioning is used to determine the provenance of the stone blades. In Europe, Neolithic 'axe factories', where thousands of ground stone axes were roughed out are known from many places, such as:

- Great Langdale, Great Britain (tuff)
- Rathlin Island, Ireland (porcellanite)
- Krzemionki, Poland (flint)
- Plancher-les-Mines, France (pelite)
- Val de'Aoste, Italy (omphacite).

Stone axes are still produced and in use today in parts of Irian Jaya, New Guinea. The Mount Hagen area was an important production centre.

From the late Neolithic/Chalcolithic onwards, axes were made of copper or copper mixed with arsenic. These axes were flat and hafted much like their stone predecessors. Axes continued to be made in this manner with the introduction of Bronze metallurgy. Eventually the hafting method changed and the flat axe developed into the ‘flanged axe,’ then palstaves, and later winged and socketed axes.

The Proto-Indo-European word for "axe" may have been pelek'u- (Greek pelekus πέλεκυς, Sanskrit parashu, see also Parashurama), but the word was probably a loan, or a Neolithic wanderwort, ultimately related to Sumerian balag, Akkadian pilaku- .

Symbolism, ritual, and folklore

At least since the late Neolithic, elaborate axes (battle-axes, T-axes, etc.) had a religious significance and probably indicated the exalted status of their owner. Certain types almost never show traces of wear; deposits of unshafted axe blades from the middle Neolithic (such as at the Somerset Levels in Britain) may have been gifts to the deities.


Minoan symbolic labrys of gold, 2nd millennium BC: many Arkalochori Axes have been found in the Arkalochori cave

In Minoan Crete, the double axe (labrys) had a special significance, used by women priests in religious ceremonies. In 1998 a labrys, complete with an elaborately embellished haft, was found at Cham-Eslen, Canton of Zug, Switzerland. The haft was 120 cm long and wrapped in ornamented birch-bark. The axe blade is 17,4 cm long and made of antigorite, mined in the Gotthard-area. The haft goes through a biconical drilled hole and is fastened by wedges of antler and by birch-tar. It belongs to the early Cortaillod culture.

In the Roman *fasces*, the axe symbolized the authority to execute and were often used as symbols for Fascist Italy under Mussolini.

In folklore, stone axes were sometimes believed to be thunderbolts and were used to guard buildings against lightning, as it was believed (mythically) that lightning never struck the same place twice. This has caused some skewing of axe distributions.

Steel axes were important in superstition as well. A thrown axe could keep off a hailstorm, sometimes an axe was placed in the crops, with the cutting edge to the skies to protect the harvest against bad weather. An upright axe buried under the sill of a house would keep off witches, while an axe under the bed would assure male offspring.

Basques, Australians and New Zealanders have developed variants of rural sports that perpetuate the traditions of log cutting with axe. The Basque variants, splitting horizontally or vertically disposed logs, are generically called *aizkolaritza* (from *aizkora*: axe).


A collection of old Australian axes

In Yorùbá mythology, the oshe (double-headed axe) symbolizes Shango, Orisha (god) of thunder and lightning. It is said to represent swift and balanced justice. Shango altars often contain a carved figure of a woman holding a gift to the god with a double-bladed axe sticking up from her head.

Parts of the axe

The axe is comprised of two primary components, the axe *head*, and the *haft*.

The *axe head* is typically bounded by the *bit* (or blade) at one end, and the *poll* (or butt) at the other, though some designs feature two bits opposite each other. The top corner of the bit where the cutting edge begins is called the *toe*, and the bottom corner is known as the *heel*. Either side of the head is called the *cheek*, which is sometimes supplemented by *lugs* where the head meets the haft, and the hole where the haft is mounted is called the *eye*. The part of the bit that descends below the rest of the axe-head is called the beard, and a *bearded axe* is an antiquated axe head with an exaggerated beard that can sometimes extend the cutting edge twice the height of the rest of the head.

The *axe haft* is sometimes called the handle. Traditionally, it was made of a resilient hardwood like hickory or ash, but modern axes often have hafts made of durable synthetic materials. Antique axes and their modern reproductions, like the tomahawk, often had a simple, straight haft with a circular cross-section that wedged onto the axe-head without the aid of wedges or pins. Modern hafts are curved for better grip and to aid in the swinging motion, and are mounted securely to the head. The *shoulder* is where the head mounts onto the haft, and this is either a long oval or rectangular cross-section of the haft that's secured to the axe head with small metal or wooden wedges. The *belly* of the haft is the longest part, where it bows in gently, and the throat is where it curves sharply down into to the short *grip*, just before end of the haft, which is known as the *knob*.

Forms of Axes

Axes designed to cut or shape wood


Splitting axe

- **Felling axe** — Cuts across the grain of wood, as in the felling of trees. In single or double bit (the bit is the cutting edge of the head) forms and many different weights, shapes, handle types and cutting geometries to match the characteristics of the material being cut.
- **Splitting Axe** — Used to split with the grain of the wood. Splitting axe bits are more wedge shaped. This shape causes the axe to rend the fibres of the wood apart, without having to cut through them, especially if the blow is delivered with a twisting action at impact.
- **Broad axe** — Used with the grain of the wood in precision splitting. Broad axe bits are chisel-shaped (one flat and one bevelled edge) facilitating more controlled work.
- **Adze** — A variation featuring a head perpendicular to that of an axe. Rather than splitting wood side-by-side, it is used to rip a level surface into a horizontal piece of wood.

Axes as weapons

Mêlée

- **Battle axe** — In its most common form, an arm-length weapon borne in one or both hands. Compared to a sword swing, it delivers more cleaving power against a smaller target area, making it more effective against armor, due to concentrating more of its weight in the axehead. However, it allows much less precision than a sword does.
- **Tomahawk** — used almost exclusively by Native Americans, its blade was originally crafted of stone. Along with the familiar war version, which could be fashioned as a throwing weapon, the pipe tomahawk was a ceremonial and diplomatic tool. A similar type of axe is the African nzappa zap. It has traditionally been a favorite of marines since Vietnam.

- **Spontoon Tomahawk** - A French trapper and Iroquois collaboration, this was an axe with a knife-like stabbing blade instead of the familiar wedged shape.
- **Valaška** — used by Slovak shepherds, it could double as a walking stick.
- **Ono** — a Japanese weapon wielded by *sōhei* warrior monks.

Pole Arm

- **Halberd** — a spearlike weapon with a hooked poll, effective against mounted cavalry.
- **Pole axe** — designed to defeat plate armour. Its axe (or hammer) head is much narrower than other axes, which accounts for its penetrating power.
- **Danish axe** — A long-handled weapon with a large flat blade, often attributed to the Vikings.

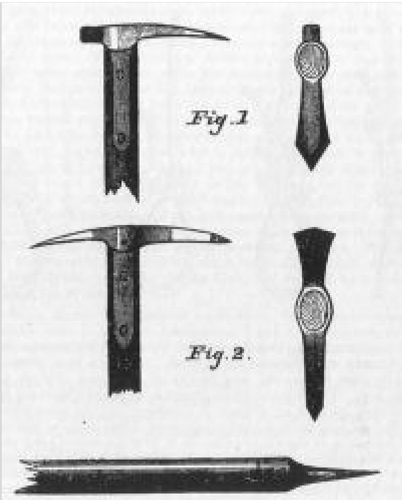
Ranged

- **Throwing axe** — Any of a number of ranged weapons designed to strike with a similar splitting action as their Mêlée counterparts. These are often small in profile and usable with one hand.
- **Hurlbat** — An entirely metal throwing axe sharpened on every auxiliary end to a point or blade, practically guaranteeing some form of damage against its target.
- **Francisca** or **Frankish axe** — a short throwing weapon of the European Migration Period, the name of which may have become attached to the Germanic tribe associated with it: the Franks (see France).

Axes for other uses

- **Firefighter's axe** or **fire axe** — It has a pick-shaped pointed poll (area of the head opposite the cutting edge). It is often decorated in vivid colors to make it easily visible during an emergency. Its primary use is for breaking down doors.
- **Pulaski** — An axe with a mattock blade built into the rear of the main axe blade, used for digging ('grubbing out') through and around roots as well as chopping. In addition to the McCloud (a tool similar to a hoe/rake combination), the pulaski is an indispensable tool used in fighting forest fires, as well as trail-building, brush clearance and similar functions.
- **Splitting maul** — A splitting implement that has evolved from the simple 'wedge' design to more complex designs. Some mauls have a conical 'axehead'; compound mauls have swiveling 'sub-wedges', among other types; others have a heavy wedge-shaped head, with a sledgehammer face opposite.
- **slater's axe** or **zax** — An axe for cutting roofing slate, with a long point on the poll for punching nail holes, and with the blade offset laterally from the handle to protect the worker's hand from flying slate chips.

Firefighter with a fire axe



Climbing axes from circa 1872

- **Climbing axe** or **ice axe** — A number of different styles of ice axe are designed for ice climbing, and, though less used today than in previous times, for rock work, especially in enlarging steps used by climbers.

In the illustration to the left, from an 1872 "Art of Travel" publication, figure 1 represents a light axe or pick which has the great advantage of lightness and handiness, with a single blade, or adze, suited to step-cutting and with a small hammer-head at the back which balances the pick, and is useful in inserting pegs into rock and ice. Figure 2 represents a travellers' axe, slightly heavier than the first, and which, at least at the time, was recommended as adapted for mountain work of all kinds.

Hammer Axe

Hammer axes (or axe-hammers) typically feature an extended poll, opposite the blade, shaped and sometimes hardened for use as a hammer. The name axe-hammer is often applied to a characteristic shape of perforated stone axe used in the Neolithic and Bronze Ages. Iron axe-hammers are found in Roman military contexts, e.g. Cramond, Edinburgh and South Shields, Tyne and Wear.

Today they are used in many different fields of work, completing all jobs from splitting wood to removal engines from vans. Tungsten is often added for weight as an upgrade, as well as six foot handles for the heavier jobs that require added force and "massive blows" such as cutting automobile frames, slicing brake rotors, rough body work, home construction, home de-construction, etc.

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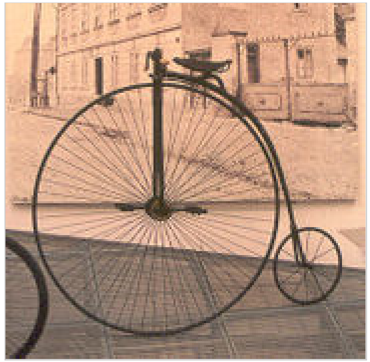
Bicycle

2008/9 Schools Wikipedia Selection. Related subjects: Road transport


A common utility bicycle


Wooden *Dandy horse* (around 1820), the first two-wheeler and as such the archetype of the bicycle

The **bicycle**, **cycle**, or **bike** is a pedal-driven, human-powered vehicle with two wheels attached to a frame, one behind the other.



A *penny-farthing* or *ordinary bicycle* photographed in the Škoda museum in the Czech Republic

Bicycles were introduced in the 19th century and now number about one billion worldwide. They are the principal means of transportation in many regions. They also provide a popular form of recreation, and have been adapted for such uses as children's toys, adult fitness, military and police applications, courier services, and competitive sports.

The basic shape and configuration of a typical bicycle has changed little since the first chain-driven model was developed around 1885. Many details have been improved, especially since the advent of modern materials and computer-aided design. These have allowed for a proliferation of specialized designs for particular types of cycling.

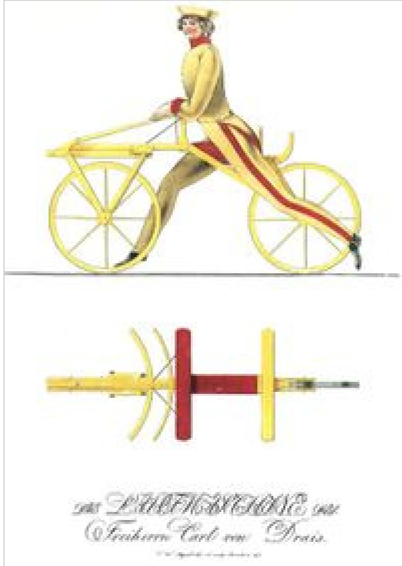
The bicycle has had a considerable effect on human society, in both the cultural and industrial realms. In its early years, bicycle construction drew on pre-existing technologies; more recently, bicycle technology has, in turn, contributed both to old and new areas.

History

Several innovators contributed to the history of the bicycle by developing precursor human-powered vehicles. The documented ancestors of today's modern bicycle were known as push bikes, Draisines or hobby horses. Being the first human means of transport to make use of the two-wheeler principle, the draisine (or *Laufmaschine*), invented by the German Baron Karl von Drais, is regarded as the archetype of the bicycle. It was introduced by Drais to the public in Mannheim in summer 1817 and in Paris in 1818. Its rider sat astride a wooden frame supported by two in-line wheels and pushed the vehicle along with his/her feet while steering the front wheel.

Evolution of the bicycle misplacing McCall's velocipede to 1830 instead of 1869

In the early 1860s, Frenchmen Pierre Michaux and Pierre Lallement took bicycle design in a new direction by adding a mechanical crank drive with pedals on an enlarged front wheel. Several why-not-the-rear-wheel inventions followed, the best known being the rod-driven velocipede by Scotsman Thomas McCall in 1869. The French creation, made of iron and wood, developed into the " penny-farthing" (more formally an *ordinary bicycle*). It featured a tubular steel frame on which were mounted wire spoked wheels with solid rubber tires. These bicycles were difficult to ride due to their very high seat and poor weight distribution.


Drais' 1817 design made-to-measure

The *dwarf ordinary* addressed some of these faults by reducing the front wheel diameter and setting the seat further back. This necessitated the addition of gearing, effected in a variety of ways, to attain sufficient speed. Having to both pedal and steer via the front wheel remained a problem. Starley's nephew, J. K. Starley, J. H. Lawson, and Shergold solved this problem by introducing the chain drive, connecting the frame-mounted pedals to the rear wheel. These models were known as *dwarf safeties*, or *safety bicycles*, for their lower seat height and better weight distribution. Starley's 1885 Rover is usually described as the first recognizably modern bicycle. Soon, the *seat tube* was added, creating the double-triangle *diamond frame* of the modern bike.


Bicycle in Plymouth, England at the start of the 20th century

Further innovations increased comfort and ushered in a second bicycle craze, the 1890s' *Golden Age of Bicycles*. In 1888, Scotsman John Boyd Dunlop introduced the pneumatic tire, which soon became universal. Soon after, the rear freewheel was developed, enabling the rider to coast. This refinement led to the 1898 invention of *coaster brakes*. Derailleur gears and hand-operated cable-pull brakes were also developed during these years, but were only slowly adopted by casual riders. By the turn of the century, cycling clubs flourished on both sides of the Atlantic, and touring and racing became widely popular.

Bicycles and horse buggies were the two mainstays of private transportation just prior to the automobile, and the grading of smooth roads in the late 19th century was stimulated by the wide use of these devices.

Uses for bicycles

Bicycles have been and are employed for many uses:

Working bicycle in Amsterdam, Netherlands.

- Utility: bicycle commuting and utility cycling
- Work: mail delivery, paramedics, police, and general delivery.
- Recreation: bicycle touring, mountain biking, BMX and physical fitness.
- Racing: track racing, criterium, roller racing and time trial to multi-stage events like the Tour of California, Giro d'Italia, the Tour de France, the Vuelta a España, the Volta a Portugal, among others.
- Military: scouting, troop movement, supply of provisions, and patrol. See bicycle infantry.
- Show: entertainment and performance, e.g. circus clowns

Transporting milk churns in Kolkata, India.

Cycling has many health benefits and does not directly contribute to global warming or environmental pollution.

Technical aspects

The bicycle has undergone continual adaptation and improvement since its inception. These innovations have continued with the advent of modern materials and computer-aided design, allowing for a proliferation of specialized bicycle types.

A half wheeler bicycle at the Golden Gate Bridge

Types of bicycle

Bicycles can be categorized in different ways: e.g. by function, by number of riders, by general construction, by gearing or by means of propulsion. The more common types include utility bicycles, mountain bicycles, racing bicycles, touring bicycles, cruiser bicycles, and BMX bicycles. Less common are tandems, lowriders, tall bikes, fixed gear, folding models and recumbents (one of which was used to set the IHPVA Hour record).

Unicycles, tricycles and quadracycles are not strictly bicycles, as they have respectively one, three and four wheels, but are often referred to informally as "bikes".

Dynamics


Bicycles leaning in a turn

A bicycle stays upright by being steered so as to keep its centre of gravity over its wheels. This steering is usually provided by the rider, but under certain conditions may be provided by the bicycle itself.

A bicycle must lean in order to turn. This lean is induced by a method known as countersteering, which can be performed by the rider turning the handlebars directly with the hands or indirectly by leaning the bicycle.

Short-wheelbase or tall bicycles, when braking, can generate enough stopping force at the front wheel in order to flip longitudinally. This action, especially if performed on purpose, is known as a stoppie, endo or front wheelie.

Performance


A racing upright bicycle

The bicycle is extraordinarily efficient in both biological and mechanical terms. The bicycle is the most efficient self-powered means of transportation in terms of energy a person must expend to travel a given distance. From a mechanical viewpoint, up to 99% of the energy delivered by the rider into the pedals is transmitted to the wheels, although the use of gearing mechanisms may reduce this by 10-15%. In terms of the ratio of cargo weight a bicycle can carry to total weight, it is also a most efficient means of cargo transportation.


A recumbent bicycle

A human being traveling on a bicycle at low to medium speeds of around 10-15 mph (15-25 km/h), using only the energy required to walk, is the most energy-efficient means of transport generally available. Air drag, which is proportional to the square of speed, requires dramatically higher power outputs as speeds increase. A bicycle which places the rider in a seated position, supine position or, more rarely, prone position, and which may be covered in an aerodynamic fairing to achieve very low air drag, is referred to as a recumbent bicycle or human powered vehicle. On an upright bicycle, the rider's body creates about 75% of the total drag of the bicycle/rider combination.

In addition, the carbon dioxide generated in the production and transportation of the food required by the bicyclist, per mile traveled, is less than 1/10th that generated by energy efficient cars.

Construction and parts

In its early years, bicycle construction drew on pre-existing technologies. More recently, bicycle technology has in turn contributed ideas in both old and new areas.

Frame

The great majority of today's bicycles have a frame with upright seating which looks much like the first chain-driven bike. Such upright bicycles almost always feature the *diamond frame*, a truss consisting of two triangles: the front triangle and the rear triangle. The front triangle consists of the head tube, top tube, down tube and seat tube. The head tube contains the headset, the set of bearings that allows the fork to turn smoothly for steering and balance. The top tube connects the head tube to the seat tube at the top, and the down tube connects the head tube to the bottom bracket. The rear triangle consists of the seat tube and paired chain stays and seat stays. The chain stays run parallel to the chain, connecting the bottom bracket to the rear dropouts. The seat stays connect the top of the seat tube (at or near the same point as the top tube) to the rear dropouts.

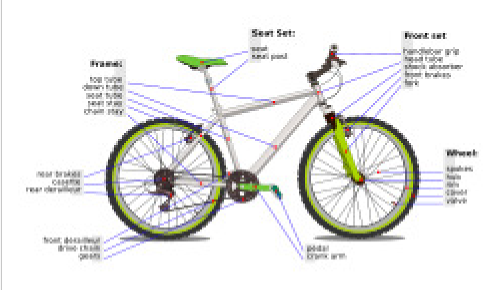


Diagram of a bicycle.


A Triumph with a step-through frame.

Historically, women's bicycle frames had a top tube that connected in the middle of the seat tube instead of the top, resulting in a lower standover height at the expense of compromised structural integrity, since this places a strong bending load in the seat tube, and bicycle frame members are typically weak in bending. This design, referred to as a *step-through frame*, allows the rider to mount and dismount in a dignified way while wearing a skirt or dress. While some women's bicycles continue to use this frame style, there is also a variation, the *mixte*, which splits the top tube into two small top tubes that bypass the seat tube and connect to the rear dropouts. The ease of stepping through is also appreciated by those with limited flexibility or other joint problems. Because of its persistent image as a "women's" bicycle, step-through frames are not common for larger frames.

Another style is the recumbent bicycle. These are inherently more aerodynamic than upright versions, as the rider may lean back onto a support and operate pedals that are on about the same level as the seat. The world's fastest bicycle is a recumbent bicycle but this type was banned from competition in 1934 by the Union Cycliste Internationale.

Historically, materials used in bicycles have followed a similar pattern as in aircraft, the goal being high strength and low weight. Since the late 1930s alloy steels have been used for frame and fork tubes in higher quality machines. Celluloid found application in mudguards, and aluminium alloys are increasingly used in components such as handlebars, seat post, and brake levers. In the 1980s aluminium alloy frames became popular, and their affordability now makes them common. More expensive carbon fibre and titanium frames are now also available, as well as advanced steel alloys and even bamboo.

Drivetrain and gearing

Since cyclists' legs are most efficient over a narrow range of pedalling speeds (cadence), a variable gear ratio helps a cyclist to maintain an optimum pedalling speed while covering varied terrain. As a first approximation, utility bicycles often use a hub gear with a small number (3 to 5) of widely-spaced gears, road bicycles and racing bicycles use derailleur gears with a moderate number (10 to 16) of closely-spaced gears, while mountain bicycles, hybrid bicycles, and touring bicycles use dérailleur gears with a larger number (15 to 30) of moderately-spaced gears, often including an extremely low gear (granny gear) for climbing steep hills.

A set of rear sprockets (also known as a cassette) and a derailleur

Different gears and ranges of gears are appropriate for different people and styles of cycling. Multi-speed bicycles allow gear selection to suit the circumstances, e.g. it may be comfortable to use a high gear when cycling downhill, a medium gear when cycling on a flat road, and a low gear when cycling uphill. In a lower gear every turn of the pedals leads to fewer rotations of the rear wheel. This allows the force required to move the same distance to be distributed over more pedal turns, reducing fatigue when riding uphill, with a heavy load, or against strong winds. A higher gear allows a cyclist to make fewer pedal cycles to maintain a given speed, but with more effort per turn of the pedals.

The *drivetrain* begins with pedals which rotate the cranks, which are held in axis by the bottom bracket. Most bicycles use a chain to transmit power to the rear wheel. A relatively small number of bicycles use a shaft drive to transmit power. A very small number of bicycles (mainly single-speed bicycles intended for short-distance commuting) use a belt drive as an oil-free way of transmitting power.

With a *chain drive* transmission, a *chainring* attached to a crank drives the chain, which in turn rotates the rear wheel via the rear sprocket(s) (cassette or freewheel). There are four gearing options: two-speed hub gear integrated with chain ring, up to 3 chain rings, up to 10 sprockets, hub gear built in to rear wheel (3-speed to 14-speed). The most common options are either a rear hub or multiple chain rings combined with multiple sprockets (other combinations of options are possible but less common).

A bicycle with shaft drive instead of a chain

With a *shaft drive* transmission, a gear set at the bottom bracket turns the shaft, which then turns the rear wheel via a gear set connected to the wheel's hub. There is some small loss of efficiency due to the two gear sets needed. The only gearing option with a shaft drive is to use a hub gear.

Steering and seating


Conventional dropdown handlebars with added aerobars

The handlebars turn the fork and the front wheel via the stem, which rotates within the headset. Three styles of handlebar are common. *Upright handlebars*, the norm in Europe and elsewhere until the 1970s, curve gently back toward the rider, offering a natural grip and comfortable upright position. *Drop handlebars* are "dropped", offering the cyclist either an aerodynamic "crouched" position or a more upright posture in which the hands grip the brake lever mounts. Mountain bikes feature a *straight handlebar* which can provide better low-speed handling due to the wider nature of the bars.


A Selle San Marco saddle designed for women

Saddles also vary with rider preference, from the cushioned ones favored by short-distance riders to narrower saddles which allow more room for leg swings. Comfort depends on riding position. With comfort bikes and hybrids the cyclist sits high over the seat, their weight directed down onto the saddle, such that a wider and more cushioned saddle is preferable. For racing bikes where the rider is bent over, weight is more evenly distributed between the handlebars and saddle, the hips are flexed, and a narrower and harder saddle is more efficient. Differing saddle designs exist for male and female cyclists, accommodating the genders' differing anatomies, although bikes typically are sold with saddles most appropriate for men.

A recumbent bicycle has a reclined chair-like seat that some riders find more comfortable than a saddle, especially riders who suffer from certain types of seat, back, neck, shoulder, or wrist pain. Recumbent bicycles may have either under-seat or over-seat steering.

Brakes

Modern bicycle *brakes* are either *rim brakes*, in which friction pads are compressed against the wheel rims, *internal hub brakes*, in which the friction pads are contained within the wheel hubs, or *disc brakes*. Disc brakes are common on off-road bicycles, tandems and recumbent bicycles.


Linear-pull brake on rear wheel of a mountain bike

A front disc brake, mounted to the fork and hub

With hand-operated brakes, force is applied to brake levers mounted on the handlebars and transmitted via Bowden cables or hydraulic lines to the friction pads. A rear hub brake may be either hand-operated or pedal-actuated, as in the back pedal *coaster brakes* which were popular in North America until the 1960s, and are still common in children's bicycles.

Track bicycles do not have brakes. Brakes are not required for riding on a track because all riders ride in the same direction around a track which does not necessitate sharp deceleration. Track riders are still able to slow down because all track bicycles are fixed-gear, meaning that there is no freewheel. Without a freewheel, coasting is impossible, so when the rear wheel is moving, the crank is moving. To slow down one may apply resistance to the pedals.

Suspension

Bicycle suspension refers to the system or systems used to *suspend* the rider and all or part of the bicycle. This serves two purposes:

This **mountain bicycle** features oversized tires, a full-suspension frame, two disc brakes and handlebars oriented perpendicular to the bike's axis

- To keep the wheels in continuous contact with rough surfaces in order to improve control.
- To isolate the rider and luggage from jarring due to rough surfaces.

Bicycle suspensions are used primarily on mountain bicycles, but are also common on hybrid bicycles, and can even be found on some road bicycles, as they can help deal with problematic vibration. Suspension is especially important on recumbent bicycles, since while an upright bicycle rider can stand on the pedals to achieve some of the benefits of suspension, a recumbent rider cannot.

Wheels

A bicycle wheel is almost always built up from a hub, rim, and spokes, and fitted with rubber pneumatic tires.

Spokes are steel or stainless steel, and can be replaced if broken. Hubs and rims can be aluminum or steel, but steel wheels are becoming rare in most countries. Aluminium rims are lighter and give much better braking in wet conditions. Typically they are anodized except for the braking surfaces. With disc brakes, the whole rim can be anodized, usually in black or silver. Wheels may also be cast or molded in one piece from aluminum alloy, plastic, and carbon fibre for various specialty bikes; plastic, for example, was once favored for BMX bikes.

The wheel axle fits into dropouts in the frame and forks. A pair of wheels may be called a wheelset, especially in the context of ready-built "off the shelf", performance-oriented wheels.

Tires vary enormously. Skinny, road-racing tires may be completely smooth, or (slick). On the opposite extreme, off-road tires are wider and thicker, and usually have a deep tread for gripping in muddy conditions.

Accessories, repairs, and tools

Some components, which are often optional accessories on sports bicycles, are standard features on utility bicycles to enhance their usefulness and comfort. Mudguards, or fenders, protect the cyclist and moving parts from spray when riding through wet areas and chainguards protect clothes from oil on the chain. Kick stands keep a bicycle upright when parked. Front-mounted baskets for carrying goods are often used. Luggage carriers and panniers can be used to carry equipment or cargo. Parents sometimes add rear-mounted child seats and/or an auxiliary saddle fitted to the crossbar to transport children.

Touring bicycle equipped with head lamp, pump, rear rack, fenders/mud-guards, water bottles and cages, and numerous saddle-bags.

Toe-clips and *toestraps* and clipless pedals help to keep the foot planted firmly in the proper position on the pedals, and enable the cyclist to pull as well as push the pedals. Technical accessories include cyclocomputers for measuring speed and distance. Other accessories include lights, reflectors, security lock, mirror, water bottles and cages, and bell.

Bicycle helmets may help reduce injury in the event of a collision or accident, and a certified helmet is legally required for some riders in some jurisdictions. Helmets are classified as an accessory or an item of clothing by others.

Many cyclists carry *tool kits*. These may include a tire patch kit (which, in turn, may contain any combination of a tire pump or CO2 cartridges, tire levers, spare tubes, self-adhesive patches, or tube-patching material, an adhesive, a piece of sandpaper or a metal grater to clean off a section of the tube, and sometimes even a block of French chalk.), wrenches, hex keys, screwdrivers, and a chain tool. There are also cycling specific multi-tools that combine many of these implements into a single compact device. More specialized bicycle components may require more complex tools, including proprietary tools specific for a given manufacturer.

Puncture repair kit with tire levers, sandpaper to clean off an area of the inner tube around the puncture, a tube of rubber solution (vulcanising fluid), round and oval patches, a metal grater and piece of chalk to make chalk powder (to dust over excess rubber solution). Kits often also include a wax crayon to mark the puncture location.

Some bicycle parts, particularly hub-based gearing systems, are complex, and many cyclists prefer to leave maintenance and repairs to professional bicycle mechanics. In some areas it is possible to purchase road-side assistance from companies such as the Better World Club. Other cyclists maintain their own bicycles, perhaps as part of their enjoyment of the hobby of cycling or simply for economic reasons.

Standards

A number of formal and industry standards exist for bicycle components, to help make spare parts exchangeable:

- ISO 5775 Bicycle tire and rim designations
- ISO 8090 Cycles—Terminology (same as BS 6102-4)
- ISO 4210 Cycles—Safety requirements for bicycles

Parts

For details on specific bicycle parts, see list of bicycle parts and category:bicycle parts.

Social and historical aspects

The bicycle has had a considerable effect on human society, in both the cultural and industrial realms.

Bicycles in daily life

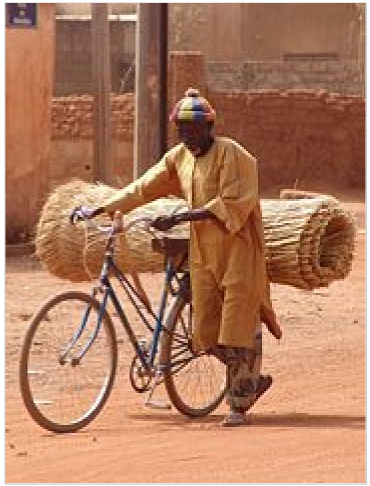
Around the turn of the 20th century, bicycles helped reduce crowding in inner-city tenements by allowing workers to commute from more spacious dwellings in the suburbs. They also reduced dependence on horses, with all the knock-on effects this brought to society. Bicycles allowed people to travel for leisure into the country, since bicycles were three times as energy efficient as walking, and three to four times as fast.

A commuting bike in Amsterdam

A bike-sharing station in Barcelona

Recently, several European cities have implemented successful schemes, known as Community bicycle programs or bike-sharing schemes. These initiatives are designed to complement a city's public transport system and offer an alternative to motorized traffic to help reduce congestion and pollution. Users can take a bicycle at a parking station, use it for a limited amount of time, and then return it to the same, or a different, station. Examples of such schemes are Bicing in Barcelona, Vélo'v in Lyon and Vélib' in Paris.

In cities where the bicycle is not an integral part of the planned transportation system, commuters often use bicycles as elements of a mixed-mode commute, where the bike is used to travel to and from train stations or other forms of rapid transit. Folding bicycles are useful in these scenarios, as they are less cumbersome when carried aboard.


A man uses a bicycle to cargo goods in Ouagadougou, Burkina Faso (2007)

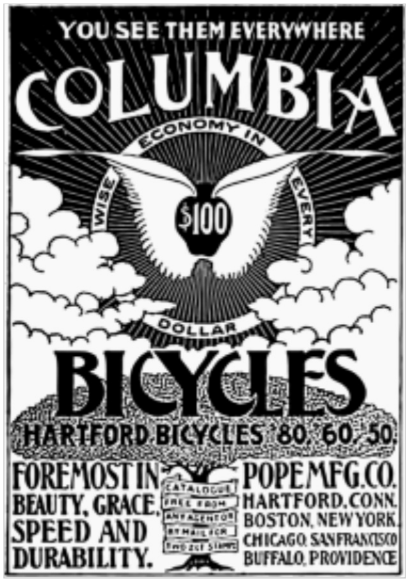
Bicycles also offer an important mode of transport in many developing contries. Until recently, bicycles have been a staple of everyday life throughout Asian countries. They are the most frequently used method of transport for commuting to work, school, shopping, and life in general. As a result bicycles there are almost always equipped with baskets and back seats.

Female emancipation


Woman with bicycle, 1890s

The diamond-frame safety bicycle gave women unprecedented mobility, contributing to their emancipation in Western nations. As bicycles became safer and cheaper, more women had access to the personal freedom they embodied, and so the bicycle came to symbolize the New Woman of the late nineteenth century, especially in Britain and the United States.

The bicycle was recognized by nineteenth-century feminists and suffragists as a "freedom machine" for women. American Susan B. Anthony said in a *New York World* interview on February 2, 1896: "Let me tell you what I think of bicycling. I think it has done more to emancipate women than anything else in the world. It gives women a feeling of freedom and self-reliance. I stand and rejoice every time I see a woman ride by on a wheel...the picture of free, untrammeled womanhood." In 1895 Frances Willard, the tightly-laced president of the Women’s Christian Temperance Union, wrote a book called *How I Learned to Ride the Bicycle*, in which she praised the bicycle she learned to ride late in life, and which she named "Gladys", for its "gladdening effect" on her health and political optimism. Willard used a cycling metaphor to urge other suffragists to action, proclaiming, "I would not waste my life in friction when it could be turned into momentum."



Columbia Bicycles advertisement from 1886

Male anger at the freedom symbolized by the New (bicycling) Woman was demonstrated when the male undergraduates of Cambridge University showed their opposition to the admission of women as full members of the university by hanging a woman bicyclist in effigy in the main town square. This was as late as 1897. The bicycle craze in the 1890s also led to a movement for so-called rational dress, which helped liberate women from corsets and ankle-length skirts and other restrictive garments, substituting the then-shocking bloomers.

Economic implications

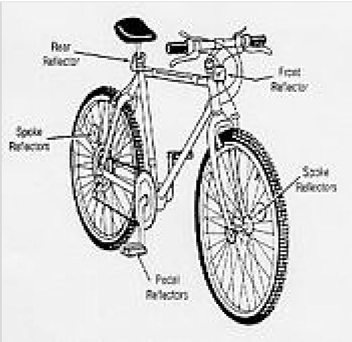
Bicycle manufacturing proved to be a training ground for other industries and led to the development of advanced metalworking techniques, both for the frames themselves and for special components such as ball bearings, washers, and sprockets. These techniques later enabled skilled metalworkers and mechanics to develop the components used in early automobiles and aircraft. J. K. Starley's company became the Rover Cycle Company Ltd. in the late 1890s, and then simply the Rover Company when it started making cars. The Morris Motor Company (in Oxford) and Škoda also began in the bicycle business, as did the Wright Brothers. Alistair Craig whose company eventually emerged to become the engine manufacturers Ailsa Craig also started from manufacturing bicycles in Glasgow in March 1885.

In general, U.S. and European cycle manufacturers used to assemble cycles from their own frames and components made by other companies, although very large companies (such as Raleigh) used to make almost every part of a bicycle (including bottom brackets, axles, etc.) In recent years, those bicycle makers have greatly changed their methods of production. Now, almost none of them produce their own frames.


Bike on beach in Goa, India

Many newer or smaller companies only design and market their products; the actual production is done by Asian companies. For example, some sixty percent of the world's bicycles are now being made in China. Despite this shift in production, as nations such as China and India become more wealthy, their own use of bicycles has declined due to the increasing affordability of cars and motorcycles. One of the major reasons for the proliferation of Chinese-made bicycles in foreign markets is the lower cost of labour in China.

Legal requirements



Reflectors for riding after dark

The 1968 Vienna Convention on Road Traffic of the United Nations considers a bicycle to be a vehicle, and a person controlling a bicycle is considered a driver. The traffic codes of many countries reflect these definitions and demand that a bicycle satisfy certain legal requirements, sometimes even including licensing, before it can be used on public roads. In many jurisdictions it is an offence to use a bicycle that is not in roadworthy condition.

In most jurisdictions, bicycles must have functioning front and rear lights when ridden after dark. As some generator or dynamo-driven lamps only operate while moving, rear reflectors are frequently also mandatory. Since a moving bicycle makes little noise, some countries insist that bicycles have a warning bell for use when approaching pedestrians, equestrians and other bicyclists.

Bicycle Brands

For a list of bicycle manufacturers see: List of bicycle manufacturing companies

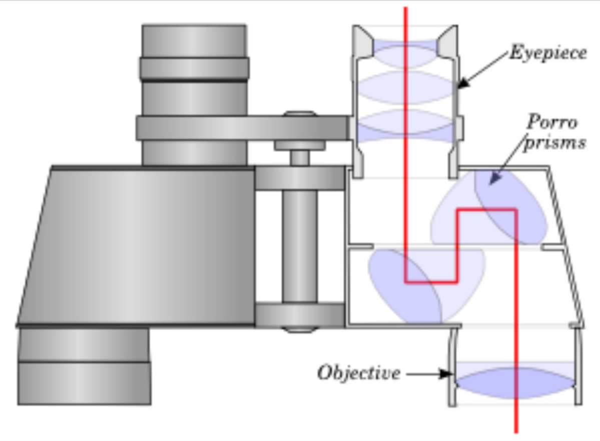
Binoculars

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Binocular telescopes, or **binoculars**, (also known as field glasses) are two identical or mirror-symmetrical telescopes mounted side-by-side and aligned to point accurately in the same direction, allowing the viewer to use both eyes (binocular vision) when viewing distant objects. Most are sized to be held using both hands, although there are much larger types.

Unlike a monocular telescope, a binocular gives users a three-dimensional image: the two views, presented from slightly different viewpoints to each of the viewer's eyes, produce a merged view with depth perception. There is no need to close or obstruct one eye to avoid confusion, as is usual with monocular telescopes.

Optical design



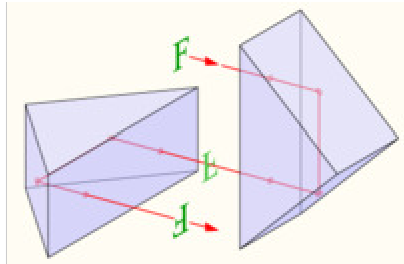
A typical Porro prism binocular design

Galilean binoculars

Almost from the invention of the telescope in the 17th century the advantages of mounting two of them side by side for binocular vision seems to have been explored. Most early binoculars used Galilean optics; that is they used a convex objective and a concave eyepiece lens. The Galilean design has the advantage of presenting an erect image but has a narrow field of view and is not capable of very high magnification. This type of construction is still used in very cheap models and in " opera glasses" or theatre glasses.

Porro prism binoculars

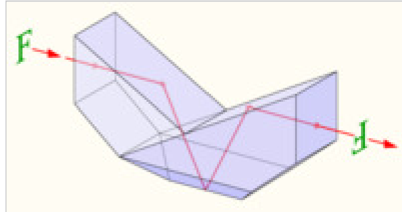

Galilean binoculars



Double Porro prism design

Named after Italian optician Ignazio Porro who patented this image erecting system in 1854 and later refined by makers like Carl Zeiss in the 1890s, binoculars of this type use a **Porro prism** in a double prism Z-shaped configuration to erect the image. This feature results in binoculars that are wide, with objective lenses that are well separated but offset from the eyepieces. Porro prism designs have the added benefit of folding the optical path so that the physical length of the binoculars is less than the focal length of the objective and wider spacing of the objectives gives better sensation of depth.

Roof prism binoculars



Abbe-Koenig "roof prism" design

Binoculars using **Roof prisms** may have appeared as early as the 1880s in a design by Achille Victor Emile Daubresse . Most roof prism binoculars use either the Abbe-Koenig prism (named after Ernst Karl Abbe and Albert Koenig and patented by Carl Zeiss in 1905) or Schmidt-Pechan prism (invented in 1899) designs to erect the image and fold the optical path. They have objective lenses that are approximately in line with the eyepieces.

Porro vs. Roof prisms

Roof-prisms designs create an instrument that is narrower and more compact than Porro prisms. There is also a difference in image brightness. Porro-prism binoculars will inherently produce a brighter image than roof-prism binoculars of the same magnification, objective size, and optical quality, because the roof-prism design employs silvered surfaces that reduce light transmission by 12% to 15%. Roof-prisms designs also require tighter tolerances as far as alignment of their optical elements (collimation). This adds to their expense since the design requires them to use fixed elements that need to be set at a high degree of collimation at the factory. Porro prisms binoculars occasionally need their prism sets to be re-aligned to bring them into collimation. The fixed alignment in roof-prism designs means the binoculars normally won't need re-collimation.

Optical parameters

Binoculars are usually designed for the specific application for which they are intended. Those different designs create certain optical parameters (some of which may be listed on the prism cover plate of the binocular). Those parameters are:

Magnification — The ratio of the focal length of the eyepiece divided into the focal length of the objective gives the linear magnifying power of binoculars (sometimes expressed as "diameters"). A magnification of factor 7, for example, produces an image as if one were 7 times closer to the object. The amount of magnification depends upon the application the binoculars are designed for. Hand-held binoculars have lower magnifications so they will be less susceptible to shaking. A larger magnification leads to a smaller field of view.

Objective diameter – The diameter of the objective lens determines how much light can be gathered to form an image. It is usually expressed in millimeters.

It is customary to categorize binoculars by the magnification × the objective diameter; e.g. 7×50.

Field of view — The field of view of a binocular is determined by its optical design. It is usually notated in a linear value, such as how many feet (meters) in width will be seen at 1,000 yards (or 1,000 m), or in an angular value of how many degrees can be viewed.

Exit pupil — Binoculars concentrate the light gathered by the objective into a beam, the exit pupil, whose diameter is the objective diameter divided by the magnifying power. For maximum effective light-gathering and brightest image, the exit pupil should equal the diameter of the fully dilated iris of the human eye— about 7 mm, reducing with age. Light gathered by a larger exit pupil is wasted. For daytime use an exit pupil of 3 mm—matching the eye's contracted pupil—is sufficient. However, a larger exit pupil makes alignment of the eye easier and avoids dark vignetting intruding from the edges.

Eye relief — Eye relief is the distance from the rear eyepiece lens to where the image is formed. It determines the distance the observer must position his or her eye behind the eyepiece in order to see an unvignetted image. The longer the focal length of the eyepiece, the greater the eye relief. Binoculars may have eye relief ranging from few millimeters to 2.5 centimeters or more. Eye relief can be particularly important for eyeglass wearers. The eye of an eyeglass wearer is typically further from the eye piece which necessitates a longer eye relief in order to still see the entire field of view. Binoculars with short eye relief can also be hard to use in instances where it is difficult to hold them steady.

Optical coatings

Since a binocular can have 16 air-to-glass surfaces, with light lost at every surface, optical coatings can significantly affect image quality. When light strikes an interface between two materials of different refractive index (e.g., at an air-glass interface), some of the light is transmitted, some reflected. In any sort of image-forming optical instrument (telescope, camera, microscope, etc.), ideally no light should be reflected; instead of forming an image, light which reaches the viewer after being reflected is distributed in the field of view, and reduces the contrast between the true image and the background. Reflection can be reduced, but not eliminated, by applying optical coatings to interfaces. Each time light enters or leaves a piece of glass; about 5% is reflected back. This "lost" light bounces around inside the binocular, making the image hazy and hard to see. Lens coatings effectively lower reflection losses, which finally results in a brighter and sharper image. For example, 8x40 binoculars with good optical coatings will yield a brighter image than uncoated 8x50 binoculars. Light can also be reflected from the interior of the instrument, but it is simple to minimize this to negligible proportions. Contrast is also improved by good coating due to the partial elimination of internal reflections.

U.S. Navy binocular

A classic lens-coating material is magnesium fluoride; it reduces reflections from 5% to 1%. Modern lens coatings consist of complex multi-layers and reflect only 0.25% or less to yield an image with maximum brightness and natural colors. For roof-prisms, anti-phase shifting coatings are sometimes used which significantly improve contrast. The presence of a coating is typically denoted on binoculars by the following terms:

- coated optics: one or more surfaces coated.
- fully coated: all air-to-glass surfaces coated. Plastic lenses, however, if used, may not be coated.
- multi-coated: one or more surfaces are multi-layer coated.
- fully multi-coated: all air-to-glass surfaces are multi-layer coated.

Phase-corrected prism coating and dielectric prism coating are recent (in 2005) effective techniques for reducing reflections.

Mechanical design

Focusing and adjustment

Binoculars to be used to view objects that are not at a fixed distance must have a focusing arrangement. Traditionally, two different arrangements have been used to provide focus. Binoculars with "independent focus" require the two telescopes to be focused independently by adjusting each eyepiece, thereby changing the distance between ocular and objective lenses. Binoculars designed for heavy field use, such as military applications, traditionally have used independent focusing. Because general users find it more convenient to focus both tubes with one adjustment action, a second type of binocular incorporates "central focusing", which involves rotation of a central focusing wheel. In addition, one of the two eyepieces can be further adjusted to compensate for differences between the viewer's eyes (usually by rotating the eyepiece in its mount). This is known as a diopter. Once this adjustment has been made for a given viewer, the binoculars can be refocused on an object at a different distance by using the focusing wheel to move both tubes together without eyepiece readjustment.

There are also "focus-free" or "fixed-focus" binoculars. They have a depth of field from a relatively large closest distance to infinity, and perform exactly the same as a focusing model of the same optical quality (or lack of it) focused on the middle distance.


Binocular with internal elements visible

Zoom binoculars, while in principle a good idea, are generally considered not to perform very well.

Most modern binoculars have hinged-telescope construction that enables the distance between eyepieces to be adjusted to accommodate viewers with different eye separation. This adjustment feature is lacking on many older binoculars.

Image stabilization

Shake can be much reduced, and higher magnifications used, with binoculars using image-stabilization technology. Parts of the instrument which change the position of the image may be held steady by powered gyroscopes or by powered mechanisms driven by gyroscopic or inertial detectors, or may be mounted in such a way as to oppose and dampen sudden movement. Stabilization may be enabled or disabled by the user as required. These techniques allow binoculars up to 20× to be hand-held, and much improve the image stability of lower-power instruments. There are some disadvantages: the image may not be quite as good as the best unstabilized binoculars when tripod-mounted, stabilized binoculars also tend to be more expensive and heavier than similarly specified non-stabilised binoculars.

Alignment

Well-collimated binoculars, when viewed through human eyes and processed by a human brain, should produce a single circular, apparently three-dimensional image, with no visible indication that one is actually viewing two distinct images from slightly different viewpoints. Departure from the ideal will cause, at best, vague discomfort and visual fatigue, but the perceived field of view will be close to circular anyway. The cinematic convention used to represent a view through binoculars as two circles partially overlapping in a figure-of-eight shape is not true to life.

Misalignment is remedied by small movements to the prisms, often by turning screws accessible without opening the binoculars, or by adjusting the position of the objective via eccentric rings built into the objective cell. Alignment is usually done by a professional although instructions for checking binoculars for collimation errors and for collimating them can be found on the Internet.

Applications

General use

Hand-held binoculars range from small 3 x 10 Galilean opera glasses, used in theaters, to glasses with 7 to 12 diameters magnification and 30 to 50 mm objectives for typical outdoor use. Porro prism models predominate although bird watchers and hunters tend to prefer, and are prepared to pay for, the lighter but more expensive roof-prism models.


People in Orchid, Florida use binoculars for birdwatching.

Many tourist attractions have installed pedestal-mounted, coin-operated binoculars to allow visitors to obtain a closer view of the attraction. In the United Kingdom, 20 pence often gives a couple of minutes of operation, and in the United States, one or two quarters gives between one-and-a-half to two-and-a-half minutes.


Coin-operated binocular

Military

Binoculars have a long history of military use. Galilean designs were widely used up to the end of the 19th century when they gave way to porro prism types. Binoculars constructed for general military use tend to be more heavily ruggedized than their civilian counterparts. They generally avoid more fragile center focus arrangements in favour of independent focus. Prism sets in military binoculars may have redundant aluminized coatings on their prism sets to guarantee they don't lose their reflective qualities if they get wet. Military binoculars of the cold war era were sometimes fitted with passive sensors that detected active IR emissions, while modern ones usually are fitted with filters blocking laser beams. Further, binoculars designed for military usage may include a stadiametric reticle in one ocular in order to facilitate range estimation.


Naval ship binocular

There are binoculars designed specifically for civilian and military use at sea. Hand held models will be 5× to 7× but with very large prism sets combined with eyepieces designed to give generous eye relief. This optical combination prevents the image vignetting or going dark when the binocular is pitching and vibrating relative to the viewer's eye. Large, high-magnification, models with large objectives are also used in fixed mountings.

Very large binocular naval rangefinders (up to 15 meters separation of the two objective lenses, weight 10 tons, for ranging World War II naval gun targets 25 km away) have been used, although late-20th century technology made this application redundant.

Astronomical

Binoculars are widely used by amateur astronomers; their wide field of view making them useful for comet and supernova seeking (giant binoculars) and general observation (portable binoculars). The Galilean moons of Jupiter, Ceres, Neptune, Pallas and Titan are invisible to the naked eye but can readily be seen with binoculars. Although technically visible unaided in pollution-free skies, Uranus and Vesta require binoculars for practical observation. 10×50 binoculars are limited to a magnitude of around +9.5, which means asteroids like Interamnia, Davida, Europa and, except under exceptional conditions Hygiea, are too faint to be seen with binoculars. Likewise too faint to be seen with binoculars are all moons except the Galileans and Titan, and the dwarf planets Pluto and Eris.

15x70 binocular.

Of particular relevance for low-light and astronomical viewing is the ratio between magnifying power and objective lens diameter. A lower magnification facilitates a larger field of view which is useful in viewing large deep sky objects such as the Milky Way, nebula, and galaxies, though the large exit pupil means some of the gathered light is wasted. The large exit pupil will also image the night sky background, effectively decreasing contrast, making the detection of faint objects more difficult except perhaps in remote locations with negligible light pollution. Binoculars specifically for most astronomical uses have higher magnification and a larger aperture objective (in the 70mm or 80mm range) because the diameter of the objective lens determines the faintest star that can be observed. These binoculars usually require some sort of mount

Much larger binoculars have been made by amateur telescope makers, essentially using two refracting or reflecting astronomical telescopes, with mixed results. A very large professional instrument, although not one that would normally be called binoculars, is the Large Binocular Telescope in Arizona, USA, which produced its "First Light" image on October 26, 2005. The LBT comprises two 8-meter reflector telescopes. While obviously not intended to be held to the eyes of a viewer, it uses two telescopes to view the same object, giving higher resolving power than a single instrument of the same light-gathering power, and allowing interferometric use.

Manufacturers

Some notable binocular manufacturers as of 2008:

Germany

- Leica GmbH – Ultravid, Duovid, Geovid: all are roof prism.
- Optolyth – Royal, ViaNova: roof prism; Alpin, Alpin Classic: porro prism.
- Zeiss GmbH – FL, Victory, Conquest: roof prism; 7×50 BGAT/T porro, 15×60 BGA/T porro, discontinued.
- Eschenbach Optik GmbH – Farlux, Trophy, Adventure, Sektor...: some are roof prism, some porro.
- Docter (the former Carl Zeiss Jena plant in Eisfeld) Nobilem 7×50, 8×56, 10×50, 15×60: porro; Docter 7×40, 8×40, 10×40: roof prism.
- Steiner GmbH – Commander, Nighthunter: porro; Predator, Wildlife: roof prism.

Austria

- Delta Optical – binoculars, riflescopes, microscopes
- Swarovski Optik – SLC, EL: roof prism; Habicht: porro prism, but to be discontinued.
- Minox
- Optolyth – Royal: Roof; Alpin: porro
- KAHLES – riflescopes, binoculars

Japan

- Canon Inc. – I.S. series: porro variants?
- Nikon Co. – High Grade series, Monarch series, RAII, Spotter series: roof prism; Prostar series, Superior E series, E series, Action EX series: porro.
- Fujinon Co. – FMTSX, FMTSX-2, MTSX series: porro.
- Kowa Co. – BD series: Roof prism.
- Pentax Co. – DCFSP/XP series: roof prism; UCF series: inverted porro; PCFV/WP/XCF series: porro.
- Olympus Co. – EXWPI series: roof prism.
- Vixen Co. – Apex/Apex Pro: roof prism; Ultima: porro*
- Zenith
- Miyauchi Co. – specializes in oversized porro binoculars.

* Also sells OEM products manufactured by the Kamakura Koki Co. Ltd. of Japan.

China

In the early 21st century, some mid-priced binoculars have become available in the internal Chinese market. A few are said to be comparable in both performance and price to those of some of the better brands, but the great majority are inferior.

- Sicong (from Xian Stateoptics) – Navigator series: roof prism; Ares series: porro.
- WDtian (from Yunnan State optics) – porro.
- Yunnan State optics – MS series: porro.

United States

- Alpen*
- Barska
- Brunton, Inc.
- Bushnell Performance Optics*
- Carson Optical
- Leupold & Stevens, Inc.*
- Simmons
- Vortex Optics
- Weaver
- William Optics
- Zen-Ray Optics – SUMMIT, Vista Series WP.

* Also sells OEM products manufactured by the KAMAKURA KOKI CO. LTD. of Japan.

Russia

- Yukon Advanced Optics
- Baigish
- Kronos
- Russian Military Binoculars – BPOc 10x42 7x30, BKFC series.

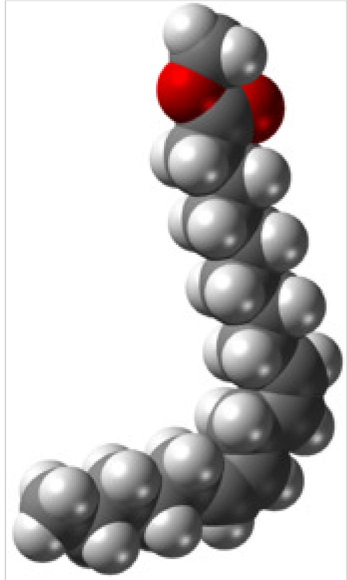
Biodiesel

2008/9 Schools Wikipedia Selection. Related subjects: Engineering; Environment

Biodiesel refers to a diesel-equivalent processed fuel consisting of short chain alkyl (methyl or ethyl) esters, made by transesterification of vegetable oils or animal fats, which can be used (alone, or blended with conventional diesel fuel) in unmodified diesel-engine vehicles.

In some countries biodiesel is less expensive than conventional diesel.

Biodiesel is distinguished from the straight vegetable oils (SVO) or waste vegetable oils (WVO) used (alone, or blended) as fuels in some diesel vehicles.



Space-filling model of Methyl Linoleate, or Linoleic Acid Methyl Ester, a common Methyl Ester produced from Soybean or Canola oil and Methanol.

On August 31, 1937, G. Chavanne of the University of Brussels (Belgium) was granted a patent for a 'Procedure for the transformation of vegetable oils for their uses as fuels' (fr. 'Procédé de Transformation d'Huiles Végétales en Vue de Leur Utilisation comme Carburants') Belgian Patent 422,877. This patent described the alcoholysis (often referred to as transesterification) of vegetable oils using ethanol (and mentions methanol) in order to separate the fatty acids from the glycerol by replacing the glycerol with short linear alcohols. This appears to be the first account of the production of what is known as 'biodiesel' today.

Biodiesel is biodegradable and non- toxic, and typically produces about 60% less net-lifecycle carbon dioxide emissions, as it is itself produced from atmospheric carbon dioxide via photosynthesis in plants. Its emissions of smog forming hydrocarbon are 65% less, although the Nitrogen Oxide emissions are about 10% greater than those from petroleum-based diesel. Net-lifetime carbon dioxide emissions can actually differ widely between fuels depending upon production methods of the source vegetable oils and processing methods employed in their creation. It is therefore debatable as to the extent that biodiesel reduces total carbon dioxide emissions currently contributing to anthropogenic global warming compared to those from petroleum-based diesel.

Some vehicle manufacturers are positive about the use of biodiesel, citing lower engine wear as one of the fuel's benefits. Biodiesel is a better solvent than standard diesel, as it 'cleans' the engine, removing deposits in the fuel lines. However, this may cause blockages in the fuel injectors if an engine has been previously run on petroleum diesel for years. For this reason, car manufacturers recommend that the fuel filter be changed a few months after switching to biodiesel (the fuel filter, as part of a routine maintenance plan, is generally replaced anyway). Most manufacturers release lists of the cars that will run on 100% biodiesel.

Space-filling model of Ethyl Stearate, or Stearic Acid Ethyl Ester, an Ethyl Ester produced from Soybean or Canola oil and Ethanol.

Other vehicle manufacturers remain cautious over use of biodiesel. In the UK many only maintain their engine warranties for use with maximum 5% biodiesel — blended in with 95% conventional diesel — although this position is generally considered to be overly cautious. Scania and Volkswagen are exceptions, allowing most of their engines to operate on 100% biodiesel. Peugeot and Citroën are also exceptions in that they have both recently announced that their PSA HDi engine can run on 30% biodiesel.

The British businessman Richard Branson's Virgin Voyager train, number 220007 *Thames Voyager* was converted to run on biodiesel, although an adverse effect occurred when it was proven to reduce reliability and to raise costs of maintenance significantly.

Biodiesel can also be used as a heating fuel in domestic and commercial boilers. Existing oil boilers may contain rubber parts and may require conversion to run on biodiesel, but the conversion process is usually relatively simple-- involving the exchanging of rubber parts for synthetic ones due to biodiesel being a strong solvent. One should not burn B100 (pure 100% biodiesel) in an existing home heater without breaking it in, as biodiesel will dissolve coagulated heating oil, which can break off in chunks and cause problems. It is suggested to start by using biodiesel as an additive, and then work your way up to burning biodiesel/petrodiesel mixes of stronger amounts. However, thanks to its strong solvent power, burning biodiesel will increase the efficiency of your home heater.

Biodiesel can be distributed using today's infrastructure, and its use and production are increasing rapidly. Fuel stations are beginning to make biodiesel available to consumers, and a growing number of transport fleets use it as an additive in their fuel. Biodiesel is generally more expensive to purchase than petroleum diesel but this differential may diminish due to economies of scale, the rising cost of petroleum and government tax subsidies.

Description

Biodiesel is a liquid which varies in colour — between golden and dark brown — depending on the production feedstock. It is practically immiscible with water, has a high boiling point and low vapor pressure. Typical methyl ester biodiesel has a flash point of ~ 150 °C (300 °F). Biodiesel has a density of ~ 0.88 g/cm³, less than that of water. Biodiesel has a viscosity similar to petrodiesel, the current industry term for diesel produced from petroleum. It can be used as an additive in formulations of diesel to increase the lubricity of pure Ultra-Low Sulfur Diesel (ULSD) fuel, which is advantageous because it has virtually no sulfur content. Much of the world uses a system known as the "B" factor to state the amount of biodiesel in any fuel mix, in contrast to the "BA" or "E" system used for ethanol mixes. For example, fuel containing 20% biodiesel is labeled B20. Pure biodiesel is referred to as B100.

Biodiesel is a renewable fuel that can be manufactured from algae, vegetable oils, animal fats or recycled restaurant greases; it can be produced locally in most countries. It is safe, biodegradable and reduces air pollutants, such as particulates, carbon monoxide and hydrocarbons. Blends of 20 percent biodiesel with 80 percent petroleum diesel (B20) can generally be used in unmodified diesel engines. Biodiesel can also be used in its pure form (B100), but may require certain engine modifications to avoid maintenance and performance problems.

The volumetric energy density of biodiesel is about 33 MJ/l. This is 9 % lower than regular Number 2 petrodiesel. Variations in biodiesel energy density is more dependent on the feedstock used than the production process. Still these variations are less than for petrodiesel. It has been claimed biodiesel gives better lubricity and more complete combustion thus increasing the engine energy output and partially compensating for the higher energy density of petrodiesel.

Historical background

Transesterification of a vegetable oil was conducted as early as 1853 by scientists E. Duffy and J. Patrick, many years before the first diesel engine became functional. Rudolf Diesel's prime model, a single 10 ft (3 m) iron cylinder with a flywheel at its base, ran on its own power for the first time in Augsburg, Germany, on August 10, 1893. In remembrance of this event, August 10 has been declared "International Biodiesel Day". Diesel later demonstrated his engine and received the *Grand Prix* (highest prize) at the World Fair in Paris, France in 1900.

This engine stood as an example of Diesel's vision because it was powered by peanut oil — a biofuel, though not *biodiesel*, since it was not transesterified. He believed that the utilization of biomass fuel was the real future of his engine. In a 1912 speech Diesel said, "the use of vegetable oils for engine fuels may seem insignificant today but such oils may become, in the course of time, as important as petroleum and the coal-tar products of the present time."

During the 1920s, diesel engine manufacturers altered their engines to utilize the lower viscosity of petrodiesel (a fossil fuel), rather than vegetable oil (a biomass fuel). The petroleum industries were able to make inroads in fuel markets because their fuel was much cheaper to produce than the biomass alternatives. The result, for many years, was a near elimination of the biomass fuel production infrastructure. Only recently, have environmental impact concerns and a decreasing price differential made biomass fuels such as biodiesel a growing alternative.

Despite the widespread use of fossil petroleum-derived diesel fuels, interest in vegetable oils as fuels in internal combustion engines is reported in several countries during the 1920's and 1930's and later during World War II. Belgium, France, Italy, the United Kingdom, Portugal, Germany, Brazil, Argentina, Japan and China have been reported to have tested and used vegetable oils as diesel fuels during this time. Some operational problems were reported due to the high viscosity of vegetable oils compared to petroleum diesel fuel, which result in poor atomization of the fuel in the fuel spray and often leads to deposits and coking of the injectors, combustion chamber and valves. Attempts to overcome these problems included heating of the vegetable oil, blending it with petroleum-derived diesel fuel or ethanol, pyrolysis and cracking of the oils.

On August 31, 1937, G. Chavanne of the University of Brussels (Belgium) was granted a patent for a "Procedure for the transformation of vegetable oils for their uses as fuels" (fr. 'Procédé de Transformation d'Huiles Végétales en Vue de Leur Utilisation comme Carburants') Belgian Patent 422,877. This patent described the alcoholysis (often referred to as transesterification) of vegetable oils using methanol and ethanol in order to separate the fatty acids from the glycerol by replacing the glycerol by short linear alcohols. This appears to be the first account of the production of what is known as "biodiesel" today.

More recently, in 1977, Brazilian scientist Expedito Parente produced biodiesel using transesterification with ethanol, and again filed a patent for the same process. This process is classified as biodiesel by international norms, conferring a "standardized identity and quality. No other proposed biofuel has been validated by the motor industry." Currently, Parente's company Tecbio is working with Boeing and NASA to certify bioquerosene (bio-kerosene), another product produced and patented by the Brazilian scientist.

Research into the use of transesterified sunflower oil, and refining it to diesel fuel standards, was initiated in South Africa in 1979. By 1983, the process for producing fuel-quality, engine-tested biodiesel was completed and published internationally. An Austrian company, Gaskoks, obtained the technology from the South African Agricultural Engineers; the company erected the first biodiesel pilot plant in November 1987, and the first industrial-scale plant in April 1989 (with a capacity of 30,000 tons of rapeseed per annum).

Throughout the 1990s, plants were opened in many European countries, including the Czech Republic, Germany and Sweden. France launched local production of biodiesel fuel (referred to as *diester*) from rapeseed oil, which is mixed into regular diesel fuel at a level of 5%, and into the diesel fuel used by some captive fleets (e.g. public transportation) at a level of 30%. Renault, Peugeot and other manufacturers have certified truck engines for use with up to that level of partial biodiesel; experiments with 50% biodiesel are underway. During the same period, nations in other parts of the world also saw local production of biodiesel starting up: by 1998, the Austrian Biofuels Institute had identified 21 countries with commercial biodiesel projects. 100% Biodiesel is now available at many normal service stations across Europe.

In September 2005 Minnesota became the first U.S. state to mandate that all diesel fuel sold in the state contain part biodiesel, requiring a content of at least 2% biodiesel.

Technical standards


Biodiesel sample

The common international standard for biodiesel is EN 14214.

There are additional national specifications. ASTM D 6751 is the most common standard referenced in the United States and Canada. In Germany, the requirements for biodiesel are fixed in the DIN EN 14214 standard and in the UK the requirements for biodiesel is fixed in the BS EN 14214 standard, although these last two standards are essentially the same as EN 14214 and are just prefixed with the respective national standards institution codes.
There are standards for three different varieties of biodiesel, which are made of different oils:

- RME (rapeseed methyl ester, according to DIN E 51606)
- PME (vegetable methyl ester, purely vegetable products, according to DIN E 51606)
- FME (fat methyl ester, vegetable and animal products, according to DIN V 51606)

The standards ensure that the following important factors in the fuel production process are satisfied:

- Complete reaction.
- Removal of glycerin.
- Removal of catalyst.
- Removal of alcohol.

- Absence of free fatty acids.
- Low sulfur content.

Basic industrial tests to determine whether the products conform to the standards typically include gas chromatography, a test that verifies only the more important of the variables above. Tests that are more complete are more expensive. Fuel meeting the quality standards is very non-toxic, with a toxicity rating (LD_{50}) of greater than 50 mL/kg.

Applications

Biodiesel can be used in pure form (B100) or may be blended with petroleum diesel at any concentration in most modern diesel engines. Biodiesel will degrade natural rubber gaskets and hoses in vehicles (mostly found in vehicles manufactured before 1992), although these tend to wear out naturally and most likely will have already been replaced with FKM, which is nonreactive to biodiesel.

Biodiesel has better lubricity than that of today's diesel fuels. During the manufacture of these, to comply with low SO_2 engine emission limits set in modern standards, severe hydrotreatment is included. Biodiesel addition reduces wear increasing the life of the fuel injection equipment that relies on the fuel for its lubrication, such as high pressure injection pumps, pump injectors (also called *unit injectors*) and fuel injectors.

Biodiesel is a better solvent than petrodiesel, and has been known to break down deposits of residue in the fuel lines of vehicles that have previously been run on petrodiesel. As a result, fuel filters and injectors may become clogged with particulates if a quick transition to pure biodiesel is made, as biodiesel “cleans” the engine in the process. Therefore, it is recommended to change the fuel filter within 600–800 miles after first switching to a biodiesel blend.

Use

Older diesel Mercedes are popular for running on biodiesel.

Pure, non-blended biodiesel can be poured straight into the tank of any diesel vehicle. As with normal diesel, low-temperature biodiesel is sold during winter months to prevent viscosity problems. Some older diesel engines still have natural rubber parts which will be affected by biodiesel, but in practice these rubber parts should have been replaced long ago. Biodiesel is used by millions of car owners in Europe (particularly Germany).

Research sponsored by petroleum producers has found petroleum diesel better for car engines than biodiesel. This has been disputed by independent bodies, including for example the Volkswagen environmental awareness division, who note that biodiesel reduces engine wear. Pure biodiesel produced 'at home' is in use by thousands of drivers who have not experienced failure, however, the fact remains that biodiesel has been widely available at gas stations for less than a decade, and will hence carry more risk than older fuels. Biodiesel sold publicly is held to high standards set by national standards bodies.

Gelling

The temperature at which pure (B100) biodiesel starts to gel varies significantly and depends upon the mix of esters and therefore the feedstock oil used to produce the biodiesel. For example, biodiesel produced from low erucic acid varieties of canola seed (RME) starts to gel at approximately −10 °C (14 °F). Biodiesel produced from tallow tends to gel at around +16 °C (68 °F). As of 2006, there are a very limited number of products that will significantly lower the gel point of straight biodiesel. A number of studies have shown that winter operation is possible with biodiesel blended with other fuel oils including #2 low sulfur diesel fuel and #1 diesel / kerosene. The exact blend depends on the operating environment: successful operations have run using a 65% LS #2, 30% K #1, and 5% bio blend. Other areas have run a 70% Low Sulfur #2, 20% Kerosene #1, and 10% bio blend or an 80% K#1, and 20% biodiesel blend. According to the National Biodiesel Board (NBB), B20 (20% biodiesel, 80% petrodiesel) does not need any treatment in addition to what is already taken with petrodiesel.

Some people modify their vehicles to permit the use of biodiesel without mixing and without the possibility of gelling at low temperatures. This practice is similar to the one used for running straight vegetable oil. They install a second fuel tank (some models of trucks have two tanks already). This second fuel tank is insulated and a heating coil using engine coolant is run through the tank. There is then a temperature sensor installed to notify the driver when the fuel is warm enough to burn, the driver then switches which tank the engine is drawing from.

Contamination by water

Biodiesel may contain small but problematic quantities of water. Although it is hydrophobic (non-miscible with water molecules), it is said to be, at the same time, hygroscopic to the point of attracting water molecules from atmospheric moisture; in addition, there may be water that is residual to processing or resulting from storage tank condensation. The presence of water is a problem because:

- Water reduces the heat of combustion of the bulk fuel. This means more smoke, harder starting, less power.
- Water causes corrosion of vital fuel system components: fuel pumps, injector pumps, fuel lines, etc.
- Water & microbes cause the paper element filters in the system to fail (rot), which in turn results in premature failure of the fuel pump due to ingestion of large particles.
- Water freezes to form ice crystals near 0 °C (32 °F). These crystals provide sites for nucleation and accelerate the gelling of the residual fuel.
- Water accelerates the growth of microbe colonies, which can plug up a fuel system. Biodiesel users who have heated fuel tanks therefore face a year-round microbe problem.

Previously, the amount of water contaminating biodiesel has been difficult to measure by taking samples, since water and oil separate. However, it is now possible to measure the water content using water in oil sensors.

Heating applications

Biodiesel can also be used as a heating fuel in domestic and commercial boilers. A technical research paper describes laboratory research and field trials project using pure biodiesel and biodiesel blends as a heating fuel in oil fired boilers. During the Biodiesel Expo 2006 in the UK, Andrew J. Robertson presented his biodiesel heating oil research from his technical paper and suggested that B20 biodiesel could reduce UK household CO_2 emissions by 1.5 million tonnes per year and would only require around 330,000 hectares of arable land for the required biodiesel for the UK heating oil sector. The paper also suggests that existing oil boilers can easily and cheaply be converted to biodiesel if B20 biodiesel is used.

Availability and prices

Global biodiesel production reached 3.8 million tons in 2005. Approximately 85% of biodiesel production came from the European Union.

Retail, at the pump, prices including Federal and state motor taxes, of B2/B5 are lower than petroleum diesel by about 12 cents, and B20 blends are the same as petrodiesel.

Production

Chemically, transesterified biodiesel comprises a mix of mono- alkyl esters of long chain fatty acids. The most common form uses methanol to produce methyl esters as it is the cheapest alcohol available, though ethanol can be used to produce an ethyl ester biodiesel and higher alcohols such as isopropanol and butanol have also been used. Using alcohols of higher molecular weights improves the cold flow properties of the resulting ester, at the cost of a less efficient transesterification reaction. A lipid transesterification production process is used to convert the base oil to the desired esters. Any Free fatty acids (FFAs) in the base oil are either converted to soap and removed from the process, or they are esterified (yielding more biodiesel) using an acidic catalyst. After this processing, unlike straight vegetable oil, biodiesel has combustion properties very similar to those of petroleum diesel, and can replace it in most current uses.

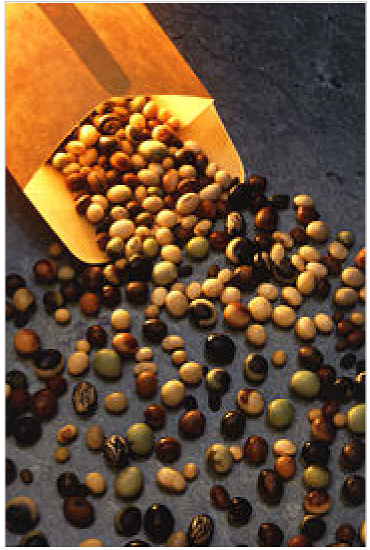
A byproduct of the transesterification process is the production of glycerol. For every 1 tonne of biodiesel that is manufactured, 100 kg of glycerol are produced. Originally, there was a valuable market for the glycerol, which assisted the economics of the process as a whole. However, with the increase in global biodiesel production, the market price for this crude glycerol (containing 20% water and catalyst residues) has crashed. Research is being conducted globally to use this glycerol as a chemical building block. One initiative in the UK is The Glycerol Challenge.

Usually this crude glycerol has to be purified, typically by performing vacuum distillation. This is rather energy intensive. The refined glycerol (98%+ purity) can then be utilised directly, or converted into other products. The following announcements were made in 2007: A joint venture of Ashland Inc. and Cargill announced plans to make propylene glycol in Europe from glycerol and Dow Chemical announced similar plans for North America. Dow also plans to build a plant in China to make epichlorhydrin from glycerol. Epichlorhydrin is a raw material for epoxy resins.

Biodiesel feedstock

A variety of oils can be used to produce biodiesel. These include:

- Virgin oil feedstock; rapeseed and soybean oils are most commonly used, soybean oil alone accounting for about ninety percent of all fuel stocks; It also can be obtained from field pennycress and Jatropha other crops such as mustard, flax, sunflower, palm oil, hemp, and even algae show promise (see List of vegetable oils for a more complete list);
- Waste vegetable oil (WVO);
- Animal fats including tallow, lard, yellow grease, chicken fat, and the by-products of the production of Omega-3 fatty acids from fish oil.
- Sewage. A company in New Zealand has successfully developed a system for using sewage waste as a substrate for algae and then producing bio-diesel.



Soybeans are used as a source of biodiesel

Plant oils

Worldwide production of vegetable oil and animal fat is not yet sufficient to replace liquid fossil fuel use. Furthermore, some environmental groups object to the vast amount of farming and the resulting over-fertilization, pesticide use, and land use conversion that they say would be needed to produce the additional vegetable oil.

Many advocates suggest that waste vegetable oil is the best source of oil to produce biodiesel. However, the available supply is drastically less than the amount of petroleum-based fuel that is burned for transportation and home heating in the world. It is important to note that one unit of waste oil is not equivalent to one unit of biodiesel.

Although it is economically profitable to use WVO to produce biodiesel, it is even more profitable to convert WVO into other products such as soap. Therefore, most WVO that is not dumped into landfills is used for these other purposes. Animal fats are similarly limited in supply, and it would not be efficient to raise animals simply for their fat. However, producing biodiesel with animal fat that would have otherwise been discarded could replace a small percentage of petroleum diesel usage. Currently, a 5-million dollar plant is being built, with the intent of producing biodiesel from some of the estimated 1.05 million tonnes (2.3 billion pounds) of chicken fat produced annually the local Tyson poultry plant, though insiders estimate a potential production of 1.1 billion litres (300 million gallons) of fuel from the chicken fat feedstock.

The estimated transportation diesel fuel and home heating oil used in the United States is about 190 billion litres (50 billion US gallons) according to the Energy Information Administration, US Department of Energy - . Waste vegetable oil and animal fats would not be enough to meet this demand. In the United States, estimated production of vegetable oil for all uses is about 11 million tonnes (24 billion pounds) or 11 billion litres (3 billion US gallons), and estimated production of animal fat is 5.3 million tonnes (12 billion pounds).

Biodiesel feedstock plants utilize photosynthesis to convert solar energy into chemical energy. The stored chemical energy is released when it is burned, therefore plants can offer a sustainable oil source for biodiesel production. Most of the carbon dioxide emitted when burning biodiesel is simply recycling that which was absorbed during plant growth, so the net production of greenhouse gases is small.

Feedstock yield efficiency per acre affects the feasibility of ramping up production to the huge industrial levels required to power a significant percentage of national or world vehicles. The highest yield feedstock for biodiesel is algae, which can produce 250 times the amount of oil per acre as soybeans.

Sunflowerseed oil

Types

Vegetable fats	(list)
Essential oil	(list)
Macerated	(list)

Uses

Drying oil - Oil paint
Cooking oil
Fuel - **Biodiesel**
Aromatherapy

Components

Saturated fat
Monounsaturated fat
Polyunsaturated fat
Trans fat

Yields of common crops

Crop	kg oil/ ha	litres oil/ ha	lbs oil/ acre	US gal/ acre
corn (maize)	145	172	129	18
cashew nut	148	176	132	19
oats	183	217	163	23
lupine	195	232	175	25
kenaf	230	273	205	29
calendula	256	305	229	33
cotton	273	325	244	35

hemp	305	363	272	39
soybean	375	446	335	48
coffee	386	459	345	49
linseed (flax)	402	478	359	51
hazelnuts	405	482	362	51
euphorbia	440	524	393	56
pumpkin seed	449	534	401	57
coriander	450	536	402	57
mustard seed	481	572	430	61
camelina	490	583	438	62
sesame	585	696	522	74
safflower	655	779	585	83
rice	696	828	622	88
tung oil tree	790	940	705	100
sunflowers	800	952	714	102
cocoa (cacao)	863	1,026	771	110
peanuts	890	1,059	795	113
opium poppy	978	1,163	873	124
rapeseed (Canola)	1,000	1,190	893	127
olives	1,019	1,212	910	129
castor beans	1,188	1,413	1,061	151
pecan nuts	1,505	1,791	1,344	191
jojoba	1,528	1,818	1,365	194
jatropha	1,590	1,892	1,420	202
macadamia nuts	1,887	2,246	1,685	240

Brazil nuts	2,010	2,392	1,795	255
avocado	2,217	2,638	1,980	282
coconut	2,260	2,689	2,018	287
oil palm	5,000	5,950	4,465	635
Chinese tallow	5,500	6,545	4,912	699
Algae (actual yield)*	6,894	7,660	6,151	819
Algae (theoretical yield)**	39,916	47,500	35,613	5,000

* Actual biomass algae yields from field trials conducted during the NREL's aquatic species program, converted using the actual oil content of the algae species grown in the specific trials. ** Algae yields are projected based on the sustainable average biomass yields of the NREL's aquatic species program, and an assumed oil content of 60%. Actual oil content was much less.

- Note: Chinese tallow (Triadica Sebifera, or Sapium sebiferum) is also known as the "Popcorn Tree" or Florida Aspen.
Source: Chinese tallow data, Mississippi State University
Source: Used with permission from the The Global Petroleum Club

Typical oil extraction from 100 kg. of oil seeds

Crop	Oil/100kg.
Castor Seed	50 kg
Copra	62 kg
Cotton Seed	13 kg
Groundnut Kernel	42 kg
Mustard	35 kg
Palm Kernel	36 kg
Palm Fruit	20 kg
Rapeseed	37 kg

Sesame	50 kg
Soybean	14 kg
Sunflower	32 kg

Source: Petroleum Club (with permission)
The energy content of biodiesel is about 90 percent that of petroleum diesel.

Efficiency and economic arguments

According to a study written by Drs. Van Dyne and Raymer for the Tennessee Valley Authority, the average US farm consumes fuel at the rate of 82 litres per hectare (8.75 US gallons per acre) of land to produce one crop. However, average crops of rapeseed produce oil at an average rate of 1,029 L/ha (110 US gal/acre), and high-yield rapeseed fields produce about 1,356 L/ha (145 US gal/acre). The ratio of input to output in these cases is roughly 1:12.5 and 1:16.5. Photosynthesis is known to have an efficiency rate of about 3-6% of total solar radiation and if the entire mass of a crop is utilized for energy production, the overall efficiency of this chain is known to be about 1%. This does not compare favorably to solar cells combined with an electric drive train, however biodiesel out-competes solar cells in cost and ease of deployment as solar cells still cost approximately US$1,000 per square meter. However, these statistics by themselves are not enough to show whether such a change makes economic sense. Additional factors must be taken into account, such as: the fuel equivalent of the energy required for processing, the yield of fuel from raw oil, the return on cultivating food, the effect biodiesel will have of food prices and the relative cost of biodiesel versus petrodiesel. A 1998 joint study by the U.S. Department of Energy (DOE) and the U.S. Department of Agriculture (USDA) traced many of the various costs involved in the production of biodiesel and found that overall, it yields 3.2 units of fuel product energy for every unit of fossil fuel energy consumed. That measure is referred to as the energy yield. A comparison to petroleum diesel, petroleum gasoline and bioethanol using the USDA numbers can be found at the Minnesota Department of Agriculture website In the comparison petroleum diesel fuel is found to have a 0.843 energy yield, along with 0.805 for petroleum gasoline, and 1.34 for bioethanol. The 1998 study used soybean oil primarily as the base oil to calculate the energy yields. Furthermore, due to the higher energy density of biodiesel, combined with the higher efficiency of the diesel engine, a unit of biodiesel produces the effective energy of 2.25 units of ethanol. Also, higher oil yielding crops could increase the energy yield of biodiesel. However not all oils retain the same heat capacities.

The debate over the energy balance of biodiesel is ongoing. Generally is **2.5**. Transitioning fully to biofuels could require immense tracts of land if traditional food crops are used (although non food crops can be utilized). The problem would be especially severe for nations with large economies, since energy consumption scales with economic output. If using only traditional food plants, most such nations do not have sufficient arable land to produce biofuel for the nation's vehicles. Nations with smaller economies (hence less energy consumption) and more arable land may be in better situations, although many regions cannot afford to divert land away from food production. For third world countries, biodiesel sources that use marginal land could make more sense, e.g. honge oil nuts grown along roads or jatropha grown along rail lines. More recent studies using a species of algae with up to 50% oil content have concluded that only 28,000 km² or 0.3% of the land area of the US could be utilized to produce enough biodiesel to replace all transportation fuel the country currently utilizes. Furthermore, otherwise unused desert land (which receives high solar radiation) could be most effective for growing the algae, and the algae could utilize farm waste and excess CO_2 from factories to help speed the growth of the algae. In tropical regions, such as Malaysia and Indonesia, oil palm is being planted at a rapid pace to supply growing biodiesel demand in Europe and other markets. It has been estimated in Germany that palm oil biodiesel has less than 1/3 the production costs of rapeseed biodiesel. The direct source of the energy content of biodiesel is solar energy captured by plants during photosynthesis. The website biodiesel.co.uk discusses the positive energy balance of biodiesel:

> When straw was left in the field, biodiesel production was strongly energy positive, yielding 1 GJ biodiesel for every 0.561 GJ of energy input (a yield/cost ratio of 1.78).
> When straw was burned as fuel and oilseed rapemeal was used as a fertilizer, the yield/cost ratio for biodiesel production was even better (3.71). In other words, for every unit of energy input to produce biodiesel, the output was 3.71 units (the difference of 2.71 units would be from solar energy).

Biodiesel is becoming of interest to companies interested in commercial scale production as well as the more usual home brew biodiesel user and the user of straight vegetable oil or waste vegetable oil in diesel engines. Homemade biodiesel processors are many and varied. The success of biodiesel homebrewing, and micro-economy-of-scale operations, continues to shatter the conventional business myth that large economy-of-scale operations are the most efficient and profitable. It is becoming increasingly apparent that small-scale, localized, low-impact energy keeps more resources and revenue within communities, reduces damage to the environment, and requires less waste management.

Environmental benefits

Environmental benefits in comparison to petroleum based fuels include:

- "At the tailpipe, biodiesel emits 4.7% *more* CO2 than petroleum diesel". However, if "biomass carbon [is] accounted for separately from fossil-derived carbon", one can conclude that biodiesel reduces emissions of carbon monoxide (CO) by approximately 50% and carbon dioxide by 78% on a net lifecycle basis because the carbon in biodiesel emissions is recycled from carbon that was in the atmosphere, rather than the carbon introduced from petroleum that was sequestered in the earth's crust.
- Biodiesel contains fewer aromatic hydrocarbons: benzofluoranthene: 56% reduction; Benzopyrenes: 71% reduction.
- Biodiesel can reduce by as much as 20% the direct (tailpipe) emission of particulates, small particles of solid combustion products, on vehicles with particulate filters, compared with low-sulfur (<50 ppm) diesel. Particulate emissions as the result of production are reduced by around 50%, compared with fossil-sourced diesel. (Beer et al, 2004).
- Biodiesel has a higher cetane rating than petrodiesel, which can improve performance and clean up emissions compared to crude petro-diesel (with cetane lower than 40).
- Biodiesel is considered readily biodegradable under ideal conditions and non-toxic. A University of Idaho study compared biodegradation rates of biodiesel, neat vegetable oils, biodiesel and petroleum diesel blends, and neat 2-D diesel fuel. Using low concentrations of the product to be degraded (10 ppm) in nutrient and sewage sludge amended solutions, they demonstrated that biodiesel degraded at the same rate as a dextrose control and 5 times as quickly as petroleum diesel over a period of 28 days, and that biodiesel blends doubled the rate of petroleum diesel degradation through co-metabolism . The same study examined soil degradation using 10 000 ppm of biodiesel and petroleum diesel, and found biodiesel degraded at twice the rate of petroleum diesel in soil. In all cases, it was determined biodiesel also degraded more completely than petroleum diesel, which produced poorly degradable

undetermined intermediates. Toxicity studies for the same project demonstrated no mortalities and few toxic effects on rats and rabbits with up to 5000 mg/kg of biodiesel. Petroleum diesel showed no mortalities at the same concentration either, however toxic effects such as hair loss and urinary discolouring were noted with concentrations of >2000 mg/l in rabbits.

- In the United States, biodiesel is the only alternative fuel to have successfully completed the Health Effects Testing requirements (Tier I and Tier II) of the Clean Air Act (1990).
- Since biodiesel is more often used in a blend with petroleum diesel, there are fewer formal studies about the effects on pure biodiesel in unmodified engines and vehicles in day-to-day use. Fuel meeting the standards and engine parts that can withstand the greater solvent properties of biodiesel is expected to--and in reported cases does--run without any additional problems than the use of petroleum diesel.
- The flash point of biodiesel (>150 °C) is significantly higher than that of petroleum diesel (64 °C) or gasoline (−45 °C). The gel point of biodiesel varies depending on the proportion of different types of esters contained. However, most biodiesel, including that made from soybean oil, has a somewhat higher gel and cloud point than petroleum diesel. In practice this often requires the heating of storage tanks, especially in cooler climates.
- Pure biodiesel (B100) can be used in any petroleum diesel engine, though it is more commonly used in lower concentrations. Some areas have mandated ultra-low sulfur petrodiesel, which reduces the natural viscosity and lubricity of the fuel due to the removal of sulfur and certain other materials. Additives are required to make ULSD properly flow in engines, making biodiesel one popular alternative. Ranges as low as 2% (B2) have been shown to restore lubricity. Many municipalities have started using 5% biodiesel (B5) in snow-removal equipment and other systems.

Environmental concerns

The locations where oil-producing plants are grown is of increasing concern to some environmentalists, one of the prime worries being that countries may clear cut large areas of tropical forest in order to grow such oil rich crops such as oil palm. In the Philippines and Indonesia such forest clearing is already underway for the production of palm oil for food. Loss of habitat on such a scale could endanger numerous species of plants and animals. A particular concern which has received considerable attention is the threat to the already-shrinking populations of orangutans on the Indonesian islands of Borneo and Sumatra, which face possible extinction.

Biodiesel and feedstock oils produced in Asia, South America and Africa are currently less expensive than those produced in Europe and North America suggesting that imports to these wealthier nations are likely to increase in future. Like all petroleum based fuels, biodiesel also requires a significant investment of energy before it arrives at petrol pumps, thus fair comparisons among fuels require full lifecycle analyses for each fuel type. The US EPA currently estimates that the use of biodiesel represents a 67% reduction in greenhouse gas emissions in comparison with petroleum based fuels. If deforestation, and monoculture farming techniques were used to grow biofuel crops, biodiesel is predicted to become a serious threat to the environment. These problems could be exacerbated as biodiesel becomes more popular unless stringent laws are introduced and enforced to control biodiesel production. Non-food energy crops and lipid rich algaes with vastly greater oil yields may also replace low-yield annual food crops such as soybeans, skirting the deforestation risk associated with widespread uptake of biodiesel.

As non-food crops also facilitate the use of degraded lands, wastewater, processed sewage, and other waste streams, the benefits of such crops go well beyond their greater yields. Moreover, select non-food crops such as jatropha and castorbean can be grown in polycultures, in non-till agricultural applications, and they scale well from the standpoint of production, storage, and processing. As such, these crops have considerable promise for small-scale farmers throughout tropical and temperate latitudes, providing a cash crop option which can also displace local demand for imported petroleum.

If burned without additives, Biodiesel (B100) is estimated to produce about 10% more nitrogen oxide NO_x tailpipe-emissions than petrodiesel. As biodiesel has a low sulfur content, NO_x emissions can be reduced through the use of catalytic converters to less than the NO_x emissions from conventional diesel engines. Moreover, as a transportation fuel, biodiesel is in its infancy in terms of additives which are capable of improving energy density, resistance to gelling, and NO_xemissions. Debate continues over NO_x, particulates, smog, and greenhouse gas emissions from biodiesel and all other new transportation fuels, biofuels in particular. Ultimately, greater clarity on the fundamental distinctions between smog and other local pollution issues vs. greenhouse gas emissions will be essential for both well founded public policy as well as well informend consumer choices. In February 2006 a Navy biodiesel expert claimed NO_x emissions in practice were actually lower than baseline. Further research is needed.

Recent advances in the use of cerium oxide, however, hold the potential to nearly eliminate NOx emissions from both petrodiesel and biodiesel, and diesel fuel additives based on cerium oxide can improve fuel consumption by 11% in unmodified diesel engines.

Current research

There is ongoing research into finding more suitable crops and improving oil yield. Using the current yields, vast amounts of land and fresh water would be needed to produce enough oil to completely replace fossil fuel usage. It would require twice the land area of the US to be devoted to soybean production, or two-thirds to be devoted to rapeseed production, to meet current US heating and transportation needs.

Specially bred mustard varieties can produce reasonably high oil yields, and have the added benefit that the meal leftover after the oil has been pressed out can act as an effective and biodegradable pesticide.

Algaculture

From 1978 to 1996, the U.S. National Renewable Energy Laboratory experimented with using algae as a biodiesel source in the " Aquatic Species Program". A self-published article by Michael Briggs, at the UNH Biodiesel Group, offers estimates for the realistic replacement of all vehicular fuel with biodiesel by utilizing algae that have a natural oil content greater than 50%, which Briggs suggests can be grown on algae ponds at wastewater treatment plants. This oil-rich algae can then be extracted from the system and processed into biodiesel, with the dried remainder further reprocessed to create ethanol.

The production of algae to harvest oil for biodiesel has not yet been undertaken on a commercial scale, but feasibility studies have been conducted to arrive at the above yield estimate. In addition to its projected high yield, algaculture — unlike crop-based biofuels — does not entail a decrease in food production, since it requires neither farmland nor fresh water. Some companies are pursuing algae bio-reactors for various purposes, including biodiesel production.

On May 11, 2006 the Aquaflow Bionomic Corporation in Marlborough, New Zealand announced that it had produced its first sample of bio-diesel fuel made from algae found in sewage ponds. Unlike previous attempts, the algae was naturally grown in pond discharge from the Marlborough District Council's sewage treatment works.

Boeing 787

2008/9 Schools Wikipedia Selection. Related subjects: Air & Sea transport

The **Boeing 787 *Dreamliner*** is a mid-sized, wide-body, twin engine jet airliner currently in production by Boeing Commercial Airplanes. It will carry between 210 and 330 passengers depending on variant and seating configuration. Boeing has stated that it will be more fuel-efficient than earlier Boeing airliners and will also be the first major airliner to use composite materials for most of its construction.

Until January 28, 2005, the 787 was known by the developmental designator **7E7**. Early released concept images depicted a radical design with highly curved surfaces. On April 26, 2005, a year after the launch of the program, the final look of the external 787 design was frozen, with a less rakish nose and a more conventional tail.

Boeing featured its first 787 in a rollout ceremony on July 8, 2007 at its assembly factory in Everett, Washington, by which time it had become the fastest-selling wide body airliner in history with nearly 600 orders. Originally scheduled to enter service in May 2008, production has been delayed and it is currently scheduled to enter into service in early 2009.

Boeing 787 *Dreamliner*

707 · 717 · 727 · 737 · 747 · 757 · 767 · 777 · **787**

Boeing 787-8 Dreamliner at roll-out ceremony

Type	Wide-body jet airliner
Manufacturer	Boeing Commercial Airplanes
Maiden flight	Targeted for the second quarter of 2008
Status	Production
Unit cost	**787-3:** US$146–151.5 million **787-8:** $157–167 million **787-9:** $189–200 million

Development

Background

In the late 1990s, Boeing began considering a replacement for the 767 when sales weakened due to the competing Airbus A330-200. As sales of the Boeing 747-400 were also slowing, the company proposed two new aircraft, which were the Sonic Cruiser and the 747X. The Sonic Cruiser would have achieved higher speeds (approximately Mach 0.98) while burning fuel at the same rate as the existing 767 or A330. The 747X, competing with the Airbus A380, would have lengthened the 747-400 and improved efficiency by using a composite supercritical wing.

Market interest for the 747X was tepid, but the Sonic Cruiser had brighter prospects. Several major airlines in the United States, including Continental, initially showed enthusiasm for the Sonic Cruiser concept, although they also expressed concerns about the operating cost. By decreasing travel time, they would be able to increase customer satisfaction and aircraft utilization.

The September 11, 2001 attacks upended the global airline market. Airlines could not justify large capital expenditures, and increased petroleum prices made them more interested in efficiency than speed. The worst-affected airlines, in the United States, were considered the most likely customers of the Sonic Cruiser. Boeing offered airlines the option of using the airframe for either higher speed or increased efficiency, but the high projected airframe costs caused demand to slacken further. Boeing canceled the 747X once Airbus launched production of the Airbus A380, and switched tracks by offering an alternative product, the 7E7.

Design phase

The replacement for the Sonic Cruiser project was dubbed the **7E7** (with a development code name of **Y2**.) The "E" was said to stand for various things, depending upon the audience. To some, it stood for "efficiency", to others it stood for "environmentally friendly". In the end, Boeing claimed it merely stood for "Eight", after the aircraft was eventually rechristened "787". A public naming competition was also held, for which out of 500,000 votes cast online the winning title was ***Dreamliner***.

On April 26, 2004, the Japanese airline All Nippon Airways became the launch customer for the 787, then still known as the 7E7, by announcing a firm order for 50 aircraft to be delivered at the end of 2008. ANA's order included 30 787-3, 290–330 seat, one-class domestic aircraft, and 20 787-8, long-haul, 210–250 seat, two-class aircraft for regional international routes such as Tokyo Narita–Beijing. The aircraft will allow ANA to open new routes to mid-sized cities not previously served, such as Denver, Montreal, and Boston. As is common for launch customers, ANA is rumored to have received a discount of 40–50% from list price.

Early concept images of the 787 included rakish cockpit windows, a dropped nose and a distinctive "shark-fin" vertical stabilizer. The final styling of the aircraft was more conservative, the fin appearing visually similar to those of aircraft currently in service. The nose and cockpit windows were also changed to a more conventional form.

Earlier proposed design configuration of the Boeing 7E7.

The 787-3 and 787-8 will be the initial variants, with the 787-9 entering service in 2010, despite industry rumors that it would be delayed as orders for the 787-3 and 787-8 sold out early production. Boeing initially priced the 787-8 variant at US$120 million, a low figure that surprised the industry. Boeing has since increased the price twice. As of 2007, the list price was $146–151.5 million for the 787-3, $157–167 million for the 787-8 and $189–200 million for the 787-9. Customer-announced orders and commitments for the 787 reached 237 aircraft during the first year of sales, with firm orders numbering 677 by the 787's premiere on July 8, 2007, and well before entry into service. This makes the 787 the fastest-selling wide body airliner ever before entry into service.

The engine pods on the 787 feature chevron edges to reduce noise.

The 787 uses the same technology proposed for the Sonic Cruiser in a more conventional configuration (see Features). Boeing claims the 787 will be at least 20% more fuel-efficient than current competing aircraft. One third of the efficiency gain will come from the engines, another third from aerodynamic improvements and the increased use of lighter weight composite materials, and the final third from advanced systems. The most notable contribution to efficiency is the electric architecture which replaces bleed air and hydraulic power with electrically powered compressors and pumps. Technology from the Sonic Cruiser and 787 will be used as part of Boeing's project to replace its entire airliner product line, an endeavor called the Yellowstone Project (of which the 787 is the first stage).

Boeing selected two engine types, the General Electric GEnx and Rolls-Royce Trent 1000, to power the 787, both placed in pods. Significantly, this leaves Pratt & Whitney, which normally has an entrant in the market, unable to offer one of its engines to 787 customers. According to UTC CEO George David, Pratt & Whitney "couldn't make the business case work for that engine." Also, according to industry sources, Boeing may have wished to use evolved versions of existing engines rather than the higher-risk option of an all-new engine from Pratt & Whitney. For the first time in commercial aviation, both engine types will have a standard interface with the aircraft, allowing any 787 to be fitted with either a GE or Rolls-Royce engine at any time. Engine interchangeability makes the 787 a more flexible asset to airlines, allowing them to change easily from one manufacturer's engine to the other's if required. The engine market for the 787 is estimated at US$40 billion over the next 25 years. The launch engine for all three current 787 variants is the Rolls-Royce Trent 1000. Airbus has offered the competing A350 powered by a development of the Rolls Royce Trent turbofan, the Trent XWB.

The launch of a new airliner can be expected to draw scathing comments from competitors, Boeing's doubt over the Airbus A380 and Airbus's mocking of the Sonic Cruiser being recent examples. The 787 is no exception, as Airbus's John Leahy attempted to refute all of Boeing's claims. Leahy openly criticized the large-scale use of composites in the 787's fuselage as being "rushed and ridiculous". Despite this criticism, Boeing built and tested the first composite section while examining the Sonic Cruiser concept nearly five years before, making the 787 a significantly refined product.

The 787 underwent extensive computer modeling and wind tunnel tests.

The 787 underwent wind tunnel testing at Boeing's Transonic Wind Tunnel, QinetiQ's five-meter wind tunnel at Farnborough, UK, and NASA Ames Research Centre's wind tunnel, as well as at the French aerodynamics research agency, ONERA. Since its inception in 2004, the 787 has had research and development costs ranging from more than $10-12 billion.

Boeing has stated that it is likely to develop a "stretched" version, the 787-10, with seating capacity between 290 and 310. This proposed model is intended to compete with the planned Airbus A350-900. The 787-10 would supersede the 777-200ER in Boeing’s current catalog and could also compete against the Airbus A330-300 and A340-300. Emirates Airlines, Qantas and Vietnam Airlines have shown interest in such a variant that would enter service in 2013. This variant has not yet been officially launched by Boeing, but Mike Bair, head of the 787 Program, has stated that "It's not a matter of if, but when we are going to do it... The 787-10 will be a stretched version of the 787-9 and sacrifice some range to add extra seat and cargo capacity." Although no date has been set, Boeing expects to build a freighter version, possibly in 10 to 15 years.

Production

After stiff competition, Boeing announced on December 16, 2003, that the 787 would be assembled in Everett, Washington. Instead of building the complete aircraft from the ground up in the traditional manner, final assembly employs just 800 to 1,200 people to join completed subassemblies and integrate systems. This is a technique that Boeing has previously used on the 737 program, which involves shipping fuselage barrel sections by rail from Spirit's Wichita, Kansas, facility to Boeing's narrowbody final assembly plant in Renton, Washington. As the major components have many components pre-installed before delivery to Everett, final assembly time is reduced to only three days. This is less than a fourth of the time traditionally needed for Boeing's final assembly process.

Boeing's Everett facility, selected as the site of 787 final assembly.

In order to speed delivery of the 787's major components, Boeing has modified three 747s purchased from Chinese and Taiwanese airlines. Called Dreamlifters, these widened airplanes can house the wings and fuselage of the 787 as well as other smaller parts.

Image:Boeing 747-400LCF 2.jpg

Fuselage barrel sections are flown to Everett on a 747 Large Cargo Freighter.

Boeing manufactures the 787's tail fin at its plant in Frederickson, Washington, the ailerons and flaps at Boeing Australia, and fairings at Boeing Canada Technology. For its entire history, Boeing has guarded its techniques for designing and mass producing commercial jetliner wings. For economic reasons, the wings are manufactured by Japanese companies in Nagoya, e.g. Mitsubishi Heavy Industries; the horizontal stabilizers are manufactured by Alenia Aeronautica in Italy; and the fuselage sections by Vought in Charleston, South Carolina, (USA), Alenia in Italy, Kawasaki Heavy Industries in Japan and Spirit AeroSystems, in Wichita, Kansas, (USA).

The passenger doors are made by Latecoere (France), and the cargo doors, access doors, and crew escape door are made by Saab (Sweden). Japanese industrial participation is very important to the project, with a 35% work share, and many of the subcontractors supported and funded by the Japanese government. On April 26, 2006, Japanese manufacturer Toray Industries and Boeing announced a production agreement involving $6 billion worth of carbon fibre. The deal is an extension of a contract signed in 2004 between the two companies and eases some concerns that Boeing might have difficulty maintaining its production goals for the 787.

From France, Messier-Dowty builds the landing gear and Thales supplies the integrated standby flight display and electrical power conversion system.

Honeywell and Rockwell-Collins provide flight control, guidance, and other avionics systems, including standard *dual* head up guidance systems. Future integration of forward-looking infrared is being considered by Flight Dynamics allowing improved visibility using thermal sensing as part of the HUD system, allowing pilots to "see" through the clouds.

Connecticut (USA)-based Hamilton Sundstrand provides power distribution and management systems for the aircraft, including manufacture and production of Generator Control Units (GCUs) as well as integration of power transfer systems that can move power from the Auxiliary Power Unit (APU) and the main engines to the necessary parts and machinery of the aircraft. Cold weather test of the APU took place in Alaska.

The first composite fuselage section rolled out in January 2005, and final external design was set in April 2005. On June 30, 2006, Boeing celebrated the start of major assembly of the first 787 at Fuji Heavy Industries' new factory in Handa, Japan, near Nagoya.

On December 6, 2006, Boeing conducted a "virtual rollout" of the 787. Unlike a traditional rollout (which occurred later), it took place without a physical airframe present. Taking computer aided design beyond the aircraft itself, Boeing modeled the manufacturing process, step-by-step and end-to-end, in software. The virtual rollout is intended to discover production issues prior to assembly of the first airframe, when they are cheaper to fix.

On January 12, 2007, first major assemblies, forward fuselage, center wing, and centre wheel well built by FHI and KHI were shipped on 747-400 LCF from Nagoya, Japan. They were delivered to Global Aeronautica in Charleston, South Carolina, on January 15.

Assembly of Section 41 of a 787 Dreamliner.

On February 15, 2007, the first production nose section (Section 41) was unveiled at Spirit AeroSystems in Wichita, Kansas. This was the first production nose section, used in the first complete flight-test 787 and represents those used in all subsequent production 787s. It encompasses the cockpit area, nose landing gear well, and the forward-most section of the passenger area. The section is oval-shaped (as is the entire fuselage) and is 21 feet (6.4 m) in height, 19 feet (5.74 m) in width and 42 feet (12.8 m) in length.

On March 14, 2007, the first production vertical tail fin was rolled out at Boeing's Composite Manufacturing Centre in Frederickson, Washington. On April 16, the first production all-composite nose-and-cockpit section was rolled out at Spirit Aerosystem's plant in Wichita, Kansas. The 747-400 LCF Dreamlifter delivered the first horizontal stabilizer manufactured by Alenia Aeronautica at its facility in Grottaglie, Italy to Everett on April 24.

On May 8, 2007, Vought rolled out completed rear Sections 47 and 48 from its factory in Charleston, SC. The sections were flown via the Dreamlifter to Everett, arriving on May 11 along with the all-composite forward section (section 41) manufactured by Spirit AeroSystems.

Mitsubishi Heavy Industries Ltd. shipped the first 787 carbon-fibre wings from its factory in Nagoya to Boeing's main assembly plant in Everett on May 15, 2007.

Three Dreamlifter 747s are used to transport 787 fuselage sections.

The Dreamlifter delivered the final major assembly, the integrated midbody fuselage, to Everett on May 16. Final assembly began on May 21 in Everett, Washington. Rolls-Royce shipped the first pair of Trent 1000 engines from their Derby, UK facilities on schedule on June 7 for installation on the Boeing 787. On June 26, 2007 LN1/ZA001 had finished major assembly and was towed to the paint hangar in the early morning.

Boeing started construction of a second 787. This one will be used for static testing and will not be flown. It will not be built with engines or horizontal stabilizers. Also, Boeing has stated reluctance in breaking the composite wing during the test, which would require an expensive cleanup afterwards.

An important milestone in the launch of the 787 was the certification of the Rolls Royce Trent 1000 engine on August 7, 2007, by both European and US regulators. The engine has seven variants and is the first engine to be certified for use on the aircraft.

On August 20, 2007, Hamilton Sundstrand stated that it had delivered its first two cabin air conditioning packs to Boeing for the initial flight-test of the 787

Dreamliner.

First flight & delivery delays

Boeing premiered the first 787 on July 8, 2007, which matches the aircraft's designation in the US-style month-day-year format (7/08/07). However, at the rollout, many of the airplane's parts were attached with non-aerospace fasteners, requiring it to be partly disassembled to replace them with flight fasteners afterwards. It is also understood that the aircraft's major systems and cockpit had not been installed, and no date for the initial 'power up' had been set.

Boeing originally planned for a first flight on August 27, but on August 10, 2007, Boeing spokeswoman Yvonne Leach said that the date might slip, citing factors including final assembly, avionics integration, and completion of software, hydraulic, electronic and other systems. The first flight also depends on the outcome of structural testing on the second plane on the assembly line.

On September 5, 2007, Boeing announced a three-month delay to the first flight, citing a shortage of fasteners and rivets as well as incomplete software. On October 10, 2007 a further three-month delay to the first flight and a six month delay to first deliveries was announced. The Company cited problems with its foreign and domestic supply chain in explaining the delay, especially the ongoing fastener shortage, the lack of documentation from overseas suppliers, and continuing delays in the flight guidance software provided by Honeywell. Less than a week later, the 787 program manager was replaced, although the delivery delays were not cited as a reason for the change.

On January 15, 2008, Boeing announced a further three month delay to the first flight of the 787 due to production issues. As of mid-January 2008, Boeing plans to deliver the first 787 to launch customer All Nippon Airways in early 2009.

After the initial six-month delay to first deliveries, Boeing had still intended to produce 109 of the 112 aircraft it originally planned to produce in 2008 and 2009, then increase production in 2010 to 10 aircraft a month. However, following the announcement of a further three-month delay, the Company has not yet produced a revised delivery schedule. Boeing is known to be talking to its suppliers about the possibility of future increases in production to up to 16 a month.

Design description

Features

The 787 features lighter-weight construction. Its materials (by weight) are: 50% composite, 20% aluminium, 15% titanium, 10% steel, 5% other. Composite materials are significantly lighter and stronger than traditional aircraft materials, making the 787 a very light aircraft for its capabilities. By volume, the 787 will be 80% composite. Each 787 contains approximately 35 tonnes of carbon fibre reinforced plastic, made with 23 tonnes of carbon fibre. Composites are used on fuselage, wings, tail, doors, and interior. Aluminium is used on outer fuselage skin, titanium used mainly on engines with steel used in various places.

The longest-range 787 variant can fly 8,000 to 8,500 nautical miles (14,800 to 15,700 km), enough to cover the Los Angeles to Bangkok or New York City to Taipei routes. It will have a cruise speed of Mach 0.85 (561 mph, 903 km/h at typical cruise altitudes).

The Dreamliner cabin is equipped with LED lighting and electronic window shades.

The 787 will seat 240 in two-class domestic configuration, with a 46-in (116.8 cm) pitch for first class and a 34-in (86.4 cm) pitch for coach class. 296 passengers can be seated in a high-density 3+2+3 coach arrangement with 36-in (91.4 cm) Business and 32-in (81.3 cm) Coach pitch. Up to 234 passengers may be seated in a three-class setup that uses 61-in (154.9 cm) pitch in First Class (2+2+2 or 2+1+2), 39-in (99 cm) pitch for Business (2+3+2 or 2+2+2) and 32-in (81.3 cm) for Coach (2+4+2). Cabin interior width is approximately 18 feet (547 cm) at armrest, and was increased by 1 inch (2.5 cm) over what was originally planned. The 787's interior cabin width is a full 15 in (38 cm) greater than that of the Airbus A330 and A340, but 5 in (13 cm) narrower than the proposed A350-800 XWB. For economy class in 2+4+2 or 3+2+3 arrangements, seat-bottom widths will be 18.5 in (47 cm), comparable to that found on the Boeing 777. For 3+3+3 seating, the seat widths would be approximately 17.2 in (43.7 cm), the same as those found on the Boeing 737. The vast majority of airlines are expected to select the 3+3+3 configuration on the 787.

The cabin windows are larger than others currently on in-service civil air transport (27 cm by 47 cm), with a higher eye level, so passengers can see the horizon, with Electrochromism-based "auto-dimming" to reduce cabin glare and maintain transparency. These are to be supplied by PPG. Light-emitting diode (LED) cabin lighting (three color) will be used instead of fluorescent tubes, allowing the aircraft to be entirely 'bulbless' and have 128 colour combinations.

A version of Ethernet— Avionics Full-Duplex Switched Ethernet (AFDX) / ARINC 664—will be used to transmit data between the flight deck and aircraft systems. The flight deck features LCD multi-function displays, all of which will use an industry standard GUI widget toolkit (*Cockpit Display System Interfaces to User Systems* / ARINC 661). The Lockheed Martin Orion spacecraft will use a glass cockpit derived from Rockwell Collins's 787 flight deck. Like other Boeing airliners, the 787 will use a yoke instead of a sidestick.

The internal pressure will be increased to the equivalent of 6000 feet (1800 m) altitude instead of the 8000 feet (2400 m) on conventional aircraft. According to Boeing, in a joint study with Oklahoma State University, this will significantly improve passenger comfort. Higher humidity in the passenger cabin is possible because of the use of composites (which do not corrode). Cabin air is provided by electrically driven compressors using no engine bleed air. An advanced cabin air-conditioning system provides better air quality: Ozone is removed from outside air; HEPA filters remove bacteria, viruses and fungi; and a gaseous filtration system removes odours, irritants and gaseous contaminants.

Bleedless turbofans imply the elimination of superheated air conduits normally used for de-icing, aircraft power, and other functions. These systems are to be replaced by an all-electrical system.

An Active Gust Alleviation system, similar to the system that Boeing built for the B-2 bomber, improves ride quality. Boeing, as part of its "Quiet Technology Demonstrator 2" project, is experimenting with several engine noise-reducing technologies for the 787. Among these are a redesigned air inlet containing sound-absorbing materials and redesigned exhaust duct covers whose rims are tipped in a toothed pattern to allow for quieter mixing of exhaust and outside air. Boeing expects these developments to make the 787 significantly quieter both inside and out.

Boeing engineers designed the 787 interior to better accommodate persons with mobility, sensory, and cognitive disabilities. For example, a 56-inch by 57-inch convertible lavatory includes a movable centre wall that allows two separate lavatories to become one large, wheelchair-accessible facility.

Technical concerns

Engine interchangeability

The two types of engines compatible with the 787 will use a standard electrical interface, potentially allowing any aircraft to be fitted with Rolls-Royce or GE engines at any time. This flexibility will allow an airline to switch from one manufacturer to the other in the event of technological developments that conform more closely to their operating profile. Boeing's goal is to make changing engine types as simple as a standard same-manufacturer replacement.

According to ILFC's Vice President of Marketing, Marty Olson, changing engine types on a 787 could take as long as 15 days and so be economically infeasible. "You'd have to take all the pylon, everything from the wing down, off" Olson said. He went on to complain that Boeing is still promoting the 24 hours change in spite of promises to alter their marketing. Current aircraft can have engines changed to those of a different manufacturer but this rarely happens due to the costs involved. Boeing's response is that the design is not yet finalized and 24 hours remains their goal.

Composite fuselage

Disassembled fuselage section of the Boeing 787 Dreamliner

The 787's all-composite fuselage makes it the first composite airliner in production. While the Boeing 777 contains 50% aluminium and 12% composites, the numbers for the new airplane are 15% aluminium, 50% composite (mostly carbon fibre reinforced plastic) and 12% titanium. Each fuselage barrel will be manufactured in one piece, and the barrel sections joined end to end to form the fuselage. This will eliminate the need for about 50,000 fasteners used in conventional airplane building. According to the manufacturer the composite is also more durable, allowing a higher cabin pressure during flight compared to aluminium. It was suggested by many that the risks of having a composite fuselage have not been fully assessed and should not be attempted. It was also added that carbon fibre, unlike metal, does not visibly show cracks and fatigue and repairing any damage done to the aircraft would not be easy. Boeing has dismissed such notions, insisting that composites have been used on wings and other passenger aircraft parts for many years and they have not been an issue. They have also stated that special defect detection procedures will be put in place to detect any potential hidden damage.

In 2006, Boeing launched the 787 GoldCare program. This is an optional, comprehensive life-cycle management service whereby aircraft in the program are routinely monitored and repaired as needed. This is the first program of its kind from Boeing: Post-sale protection programs are not new, but have usually been offered by third party service centers. Boeing is also designing and testing composite hardware so inspections are mainly visual. This will reduce the need for ultrasonic and other non-visual inspection methods, saving time and money.

According to Boeing Vice President Jeff Hawk, who heads the effort to certify the 787 for airline service, a crash test involving a vertical drop of a partial fuselage section from about 15 feet onto a one inch-thick steel plate went ahead as planned August 23, 2007 in Mesa, Arizona. Boeing spokeswoman Lori Gunter stated on September 6, 2007 that results matched what Boeing's engineers had predicted. As a result the company can model various crash scenarios using computational analysis rather than performing more tests on actual pieces of the plane. However, it has also been suggested by a fired Boeing engineer that in the event of a crash landing, survivable in a metal plane, the composite fuselage could shatter and burn with toxic fumes.

Weight issues

Boeing had been working to trim excess weight since assembly of the first airframe began in 2006. This is typical for new aircraft during their development phase. The aircraft is first designed on computers and an empty weight is promised to customers to ensure fuel efficiency and payload obligations. However, upon assembly, some parts may be manufactured with minor variances that multiply dramatically if the part is used frequently.

The first six 787s, which are to be used as part of the test program, will be overweight according to Boeing Commercial Airplanes CEO Scott Carson. After the flight test program, these aircraft will be delivered to airline customers All Nippon Airways, Northwest Airlines and Royal Air Maroc at speculated deeper than usual discounts. The first 787 is expected to be 5,000 lb (2,270 kg) overweight. The seventh and subsequent aircraft will be the first optimized 787s and are expected to meet all goals. Boeing has redesigned some parts and made more use of titanium. According to ILFC's Steven Udvar-Hazy, the 787-9's operating empty weight is around 14,000 lb (6,350 kg) overweight, which also could be a problem for the proposed 787-10.

Computer network vulnerability

In January 2008, Federal Aviation Administration expressed concerns about insufficient protection of the Airplane network from possible intentional or unintentional passenger access. The computer network in the passenger compartment, designed to give passengers in-flight internet access, is connected to the plane's control, navigation and communication systems.

Boeing says that although the networks were connected, various hardware and software solutions were employed to protect the plane systems such as:

- Air gaps for the physical separation of the networks,
- Firewalls, for their software separation.

However, they still have to demonstrate to FAA that they have tackled the issue.

Variants

There are three variants of the 787 and all launched at the same time in 2004. The 787-8 will enter service in 2008. The 787-3 will enter service next in 2010. The last to enter service will be the 787-9 in 2010.

787-3

This will be a 223-seat (three-class) or 296-seat (two-class) short-range version targeted at high-density flights, with a range of 2,500 to 3,050 nautical miles (4,650 to 5,650 km) when fully loaded. It is designed to compete with the Airbus A300 and to replace the Boeing 757-300 and Boeing 767-200. The 787-3's intended entry into service is 2010. This model is limited in its range not by fuel capacity but by a low maximum take-off weight of 360,000 lb (163,290 kg). It currently has the same fuel capacity as the 787-8. Actual range is calculated by the remaining available weight capacity for the fuel after the aircraft weight and payload are subtracted from the Maximum Take-Off Weight (MTOW). With a full load of passengers and cargo, it will be limited by the amount of fuel it can take on board. This is an advantage on shorter, high-density routes, especially those separated by water such as Tokyo to Shanghai, Osaka to Seoul, or London to Berlin. Many airports charge landing fees depending on the weight of the aircraft; thus, an airliner rated at a lower MTOW would pay lower fees.

The 787-3's wing will be different from the other versions. It will use large winglets instead of raked wingtips used on other variants. This decreases the wingspan and the aircraft weight. The wing will be the same except for the last 13 feet outboard of the ailerons. Due to the decrease in weight, winglets provide better efficiency over short distances than raked wingtips. The 787-3, having a MTOW considerably lower than the -8 variant, does not need the lift the extra span provides. A lower-than-optimal wing loading decreases efficiency. The twenty-six-feet wingspan reduction also enables the 787-3 to use gates reserved for medium-sized planes.

Thirty years ago it was common to fly from New York to Los Angeles in a 747. Typically, four or five flights occurred per day, one per major airline. With deregulation, more and more airlines joined the route, and existing airlines added flights to provide more variety in departure times. Overcapacity led to airlines using ever-smaller planes. Now there are around 47 direct flights per day, mostly on 737 or Airbus A320 or reconfigured 757s and 767-200s between the two cities, with each flight usually carrying around 100-160 passengers. This has led to higher congestion.

This same phenomenon is occurring in Asia, Europe, and South America. With the proliferation of open sky agreements, numerous airlines have been started in countries like Brazil, India, China, and throughout Europe. These start-ups have placed more pressure on capacity on trunk routes and have encouraged the usage of ever-smaller planes between very large cities. Routes like São Paulo to Buenos Aires, Berlin to Paris, and Mumbai to Calcutta are now increasingly being served by single-aisle planes when larger ones would reduce air traffic congestion.

Boeing believes that the pendulum has swung too far and the future of aviation between very large (but close) cities of five million or more would stabilize around the capacity level of the 787-3. It also believes legacy carriers that want to battle with low-cost airlines can use this plane with twice the capacity of a single-aisle craft but less than twice its operating cost (fuel, landing fees, maintenance, number of flight crew, airspace fees, parking fees, gate fees, etc.). Boeing sees the 787 family as a game-changer, with this variant as the most distinctive of the three.

Regions such as India and East Asia, where large population centers are in close proximity, can make good use of the 787-3. Approximately 3.1 billion people live within the range of the 787-3 if used in India or China. The 787-3's efficiency may offset the higher landing fees and acquisition costs (compared to a single-aisle plane) and make it useful on such routes. A 3,050 nm (5,600 km) range, or flight time of roughly six hours, is sufficient to connect many major cities and improve comfort and efficiency. The 787-3 is not a replacement for single-aisle planes, but it can relegate them to cities with fewer than 2 million people.

To date, however, only Japanese airlines have ordered this model for routes within East Asia.

787-8

The external appearance of the 787-8 is similar to the 787-3 with the exceptions of longer wings with raked wingtips for long-range efficiency and a substantially longer range of 7,650 to 8,200 nautical miles (14,200 to 15,200 km). The 787-8 has a substantially higher maximum take off weight than the 787-3. The 787-8 seats 223 passengers in a three class configuration. The variant will be the first of the 787 line to enter service in 2008. Boeing is targeting the 787-8 to replace the 767-200ER and 767-300ER.

787-9

This will be a "stretched" variant, seating 259 in three classes with a range of 8,000 to 8,500 nautical miles (14,800 to 15,750 km). The targeted entry into service (EIS) is set for 2010. Boeing is targeting the 787-9 to compete with Airbus's A330-200 and A340-200 and to replace their own 767-400ER. This variant is identical to the 787-8 except for structural strengthening, the lengthened fuselage, a higher fuel capacity, and a slightly wider wingspan compared to the 787-8. Each wingtip has been further extended by three feet.

When first launched, the 787-9 had the same fuel capacity as the other two variants. The design differences meant higher weight and resulted in a slightly shorter range than the 787-8. After further consultation with airlines, design changes were incorporated to add a forward tank to increase its fuel capacity. It will now have a longer range and a higher maximum take-off weight (MTOW) than the other two variants. The -9 will be able to fly non-stop from New York to Manila or from Moscow to São Paulo and will have the lowest seat-mile cost of the three 787 variants.

Early sales of this variant were limited by its 2010 entry into service. The 787-8's smaller size and earlier entry date were attractive to most airlines and led to the 787-8 receiving the most orders. With the first four years of production completely sold out, airlines have begun weighing the option of the 787-8 against the 787-9 since either one can be delivered after 2010. Analysts speculate that the 787-9's higher capacity and longer range will eventually make it the most popular variant especially among blue-chip airlines. Air New Zealand is the launch customer for the 787-9 and the second customer ever for the Boeing 787 behind ANA. Qantas, Singapore Airlines, and Continental Airlines have placed the largest orders for the 787-9.

Orders and deliveries

The Boeing 787 has not yet entered service. The first 787 is scheduled to enter passenger service in early 2009, with ANA.

Specifications

Model	787-3	787-8	787-9
Cockpit crew	Two		
Passengers	290–330	210–250	250–290
Length	186 ft (57 m)		206 ft (63 m)
Wingspan	170 ft (52 m)	197 ft (60 m)	208 ft (63 m)

Wing sweepback	32.2°		
Height	55 ft 6 in (16.92 m)		
Fuselage height	19 ft 5 in (5.91 m)		
Fuselage width	18 ft 11 in (5.75 m)		
Cabin width	18 ft (5.49 m)		
Cargo capacity	4,400 ft^3 (124.6 m^3) 28 LD3		5,400 ft^3 (152.9 m^3) 36 LD3
Empty weight	223,000 lb (101,200 kg)	242,000 lb (110,000 kg)	254,000 lb (115,200 kg)
Maximum takeoff weight	364,000 lb (165,100 kg)	484,000 lb (219,540 kg)	540,000 lb (244,940 kg)
Cruise speed	0.85 Mach (561 mph, 487 knots, 902 km/h at 40,000 ft)		
Maximum cruise speed	0.89 Mach (587 mph, 510 knots, 945 km/h at 40,000 ft)		
Range, fully loaded	2,500 – 3,050 NM (4,650 – 5,650 km)	7,650 – 8,200 NM (14,200 – 15,200 km)	8,000 – 8,500 NM (14,800 – 15,750 km)
Maximum fuel capacity	33,528 US gal (126,917 L)		36,693 US gal (138,898 L)
Service ceiling	43,000 ft (13,100 m)		
Engines (2×)	General Electric GEnx *or* Rolls-Royce Trent 1000		
Maximum thrust capability	53,000 lbf (236 kN)	64,000 lbf (285 kN)	70,000 lbf (311 kN)

Sources: Airport report, 787-3 fact sheet, 787-8 fact sheet, 787-9 fact sheet

Bridge

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

A **bridge** is a structure built to span a gorge, valley, road, railroad track, river, body of water, or any other physical obstacle, for the purpose of providing passage over the obstacle. Designs of bridges will vary depending on the function of the bridge and the nature of the terrain where the bridge is to be constructed.


The masonry Bridge of 33 Arches over the Zayandeh River is the epitome of Safavid dynasty (1502-1722) bridge design. Esfahan, Iran.


The Story Bridge in Brisbane, Australia, built in 1940 with modern materials, is an example of a steel cantilever bridge.


The Salginatobel Bridge in the Swiss Alps was declared a Historic Civil Engineering Landmark in 1991.

History

The first bridges were made by nature — as simple as a log fallen across a stream. The first bridges made by humans were probably spans of wooden logs or planks and eventually stones, using a simple support and crossbeam arrangement. Most of these early bridges could not support heavy weights or withstand strong currents. It was these inadequacies which led to the development of better bridges.


A log bridge in the French Alps near Vallorcine.

Epic literature of India provides mythological accounts of bridges constructed from India to Lanka by the army of Rama. The *Arthashastra* of Kautilya mentions the construction of dams and bridges. A Mauryan bridge near Girnar was surveyed by James Princep. The bridge was swept away during a flood, and later repaired by Puspagupta, the chief architect of emperor Chandragupta I. The bridge also fell under the care of the Yavana Tushaspa, and the Satrap Rudra Daman. The use of stronger bridges using plaited bamboo and iron chain was visible in India by about the 4th century. A number of bridges, both for military and commercial purposes, were constructed by the Mughal administration in India.


An English 18th century example of a bridge in the Palladian style, with shops on the span: Pulteney Bridge, Bath

The ancient Romans built arch bridges and aqueducts that could stand in conditions that would damage or destroy earlier designs. Some of them still stand today. An example is the Alcántara Bridge, built over the river Tagus, in Spain. Most earlier bridges would have been swept away by the strong current. The Romans also used cement, which reduced the variation of strength found in natural stone. One type of cement, called pozzolana, consisted of water, lime, sand, and volcanic rock. Brick and mortar bridges were built after the Roman era, as the technology for cement was lost then later rediscovered.

Although large Chinese bridges of wooden construction existed at the time of the Warring States, the oldest surviving stone bridge in China is the Zhaozhou Bridge, built from 595 to 605 AD during the Sui Dynasty. This bridge is also historically significant as it is the world's oldest open- spandrel stone segmental arch bridge. European segmental arch bridges date back to at least the Alconétar Bridge (approximately 2nd century AD), while the enormous Roman era Trajan's Bridge (105 AD) featured open-spandrel segmental arches in wooden construction.

Rope bridges, a simple type of suspension bridge, were used by the Inca civilization in the Andes mountains of South America, just prior to European colonization in the 1500s.

During the 18th century there were many innovations in the design of timber bridges by Hans Ulrich, Johannes Grubenmann, and others. The first book on bridge engineering was written by Hubert Gautier in 1716.

With the Industrial Revolution in the 19th century, truss systems of wrought iron were developed for larger bridges, but iron did not have the tensile strength to support large loads. With the advent of steel, which has a high tensile strength, much larger bridges were built, many using the ideas of Gustave Eiffel.

Etymology

The Oxford English Dictionary traces the origin of the word *bridge* to an Old English word *brycg*, of the same meaning, derived from a hypothetical Proto-

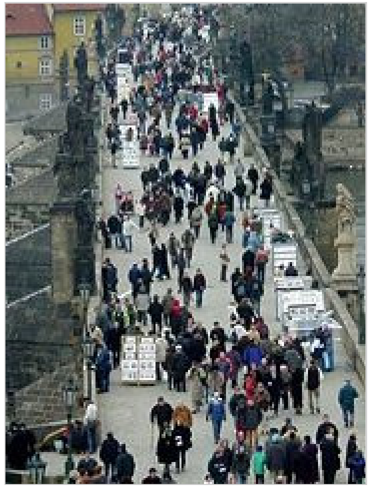
Germanic root *brugjō*. There are cognates in other Germanic languages (for instance *Brücke* in German, *brug* in Dutch, *brúgv* in Faroese or *bro* in Danish, Norwegian and Swedish).

Another theory suggests that "bridge" comes from Turkish "köprü" (lit. bridge). It is highly possible that Turkish lent this word to Eastern European languages and then, in time, it arrived in English. "Köprü" itself is derived from "köprük (köpük)" which literally means "foam".

The word for the Pope, pontiff, comes from the Latin word *pontifex* meaning "bridge builder" or simply "builder". The word "Pope" however comes from "papa" meaning "father".

Types of bridges

There are six main types of bridges: beam bridges, cantilever bridges, arch bridges, suspension bridges, cable-stayed bridges and truss bridges.



Charles Bridge in Prague

Beam bridges

Beam bridges are horizontal beams supported at each end by piers. The earliest beam bridges were simple logs that sat across streams and similar simple structures. In modern times, beam bridges are large box steel girder bridges. Weight on top of the beam pushes straight down on the piers at either end of the bridge.

Cantilever bridges

Cantilever bridges are built using cantilevers — horizontal beams that are supported on only one end. Most cantilever bridges use two cantilever arms extending from opposite sides of the obstacle to be crossed, meeting at the centre. The largest cantilever bridge is the 549 feet (167 m) Quebec Bridge in Quebec, Canada.

Arch bridges

Arch bridges are arch-shaped and have abutments at each end. The earliest known arch bridges were built by the Greeks and include the Arkadiko Bridge. The weight of the bridge is thrusted into the abutments at either side. Dubai in the United Arab Emirates is currently building the largest arch bridge in the world, which is scheduled for completion in 2012.

Suspension bridges

Suspension bridges are suspended from cables. The earliest suspension bridges were made of ropes or vines covered with pieces of bamboo. In modern bridges, the cables hang from towers that are attached to caissons or cofferdams. The caissons or cofferdams are implanted deep into the floor of a lake or river. The longest suspension bridge in the world is the 12,826 feet (3,909 m) Akashi Kaikyo Bridge in Japan.

Cable-stayed bridges

Like suspension bridges, cable-stayed bridges are held up by cables. However, in a cable-stayed bridge, less cable is required and the towers holding the cables are proportionately shorter. The first known cable-stayed bridge was designed in 1784 by C.T. Loescher. The longest cable-stayed bridge is the Tatara Bridge in the Seto Inland Sea, Japan.

Truss bridges

Truss bridges are composed of connected elements. They have a solid deck and a lattice of pin-jointed girders for the sides. Early truss bridges were made of wood, but modern truss bridges are made of metals such as wrought iron and steel. The Quebec Bridge, mentioned above as a cantilever bridge, is also the world's longest truss bridge.

By use

A bridge is designed for trains, pedestrian or road traffic, a pipeline or waterway for water transport or barge traffic. An aqueduct is a bridge that carries water, resembling a viaduct, which is a bridge that connects points of equal height. A road-rail bridge carries both road and rail traffic.

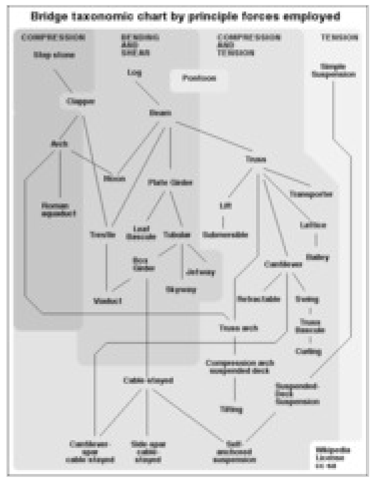
Bridges are subject to unplanned uses as well. The areas underneath some bridges have become makeshift shelters and homes to homeless people, and the undersides of bridges all around the world are spots of prevalent graffiti. Some bridges attract people attempting suicide, and become known as suicide bridges.

Decorative or ceremonial

To create a beautiful image, some bridges are built much taller than necessary. This type, often found in east-Asian style gardens, is called a Moon bridge, evoking a rising full moon. Other garden bridges may cross only a dry bed of stream washed pebbles, intended only to convey an impression of a stream. Often in palaces a bridge will be built over an artificial waterway as symbolic of a passage to an important place or state of mind. A set of five bridges cross a sinuous waterway in an important courtyard of the Forbidden City in Beijing, the People's Republic of China. The central bridge was reserved exclusively for the use of the Emperor, Empress, and their attendants.

The differences & similarities in bridge structure

Bridges may be classified by how the forces of tension, compression, bending, torsion and shear are distributed through their structure. Most bridges will employ all of the principal forces to some degree, but only a few will predominate. The separation of forces may be quite clear. In a suspension or cable-stayed span, the elements in tension are distinct in shape and placement. In other cases the forces may be distributed among a large number of members, as in a truss, or not clearly discernible to a casual observer as in a box beam. Bridges can also be classified by their lineage, which is shown as the vertical axis on the diagram to the right.



A bridge taxonomy showing evolutionary relationships

Efficiency

A bridge's *structural efficiency* may be considered to be the ratio of load carried to bridge mass, given a specific set of material types. In one common challenge students are divided into groups and given a quantity of wood sticks, a distance to span, and glue, and then asked to construct a bridge that will be tested to destruction by the progressive addition of load at the centre of the span. The bridge taking the greatest load is by this test the most *structurally efficient*. A more refined measure for this exercise is to weigh the completed bridge rather than measure against a fixed quantity of materials provided and determine the multiple of this weight that the bridge can carry, a test that emphasizes economy of materials and efficient glue joints (see *balsa wood bridge*).

A bridge's *economic efficiency* will be site and traffic dependent, the ratio of savings by having a bridge (instead of, for example, a ferry, or a longer road route) compared to its cost. The lifetime cost is composed of materials, labor, machinery, engineering, cost of money, insurance, maintenance, refurbishment, and ultimately, demolition and associated disposal, recycling, and replacement, less the value of scrap and reuse of components. Bridges employing only compression are relatively inefficient structurally, but may be highly cost efficient where suitable materials are available near the site and the cost of labor is low. For medium spans, trusses or box beams are usually most economical, while in some cases, the appearance of the bridge may be more important than its cost efficiency. The longest spans usually require suspension bridges.

Double-decker bridge

Double-decker bridges have two levels, such as the San Francisco-Oakland Bay Bridge, with two road levels. Tsing Ma Bridge and Kap Shui Mun Bridge in Hong Kong have six lanes on their upper decks, and on their lower decks there are two lanes and a pair of tracks for MTR metro trains. Some double-decker bridges only use one level for street traffic; the Washington Avenue Bridge in Minneapolis reserves its lower level for automobile traffic and its upper level for pedestrian and bicycle traffic (predominantly students at the University of Minnesota).

"Metrobridge" Vorobyovy Gory (ru:Метромост) double-deck bridge in Moscow carries the Moscow Metro.

Robert Stephenson's High Level Bridge across the River Tyne in Newcastle upon Tyne, completed in 1849, is an early example of a double-deck bridge. The upper level carries a railway, and the lower level is used for road traffic.

Another example is Craigavon Bridge in Derry, Northern Ireland. The Oresund Bridge between Copenhagen and Malmö consists of a four-lane highway on the upper level and a pair of railway tracks at the lower level.

The George Washington Bridge between New Jersey and New York has two roadway levels. It was built with only the upper roadway as traffic demands did not require more capacity. A truss work between the roadway levels provides stiffness to the roadways and reduced movement of the upper level when installed.

Tower Bridge is different example of a double-decker bridge, with the central section consisting of a low level bascule span and a high level footbridge.

More than just a bridge

- Some bridges carry special installations such as the tower of Nový Most bridge in Bratislava which carries a restaurant. Other suspension bridge towers carry transmission antennas.
- A bridge can carry overhead power lines as does the Storstrøm Bridge.
- Costs and cost overruns in bridge construction have been studied by Flyvbjerg et al. (2003). The average cost overrun in building a bridge was found to be 34%.

Aracaju-Barra Bridge in Sergipe state, Brazil.

Visual index

Index to types

Index to related topics

Politics

Most bridges are built for economic reasons, meaning they are cost justified. As the size and cost go up it becomes more difficult to cost justify bridges. Some of the largest and most costly bridges in the world are built for political reasons, namely to tie together otherwise isolated areas.

Civil engineering

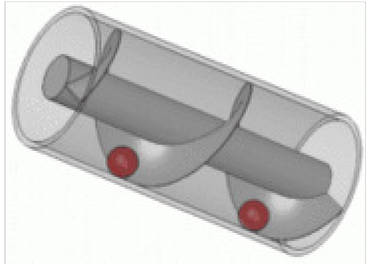
2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Civil engineering is a professional engineering discipline that deals with the design, construction and maintenance of the physical and natural built environment, including works such as bridges, roads, canals, dams and buildings. Civil engineering is the oldest engineering discipline after military engineering, and it was defined to distinguish it from military engineering. It is traditionally broken into several sub-disciplines including environmental engineering, geotechnical engineering, structural engineering, transportation engineering, water resources engineering, materials engineering, coastal engineering, surveying, and construction engineering. Civil engineering takes place on all levels: in the public sector from municipal through to federal levels, and in the private sector from individual homeowners through to international companies.

The Petronas Twin Towers, designed by Thornton-Tomasetti and Ranhill Bersekutu Sdn Bhd engineers, and Cesar Pelli, were the world's tallest buildings from 1998 to 2004.

History of the civil engineering profession

Engineering has been an aspect of life since the beginnings of human existence. Civil engineering might be considered properly commencing between 4000 and 2000 BC in Ancient Egypt and Mesopotamia when humans started to abandon a nomadic existence, thus causing a need for the construction of shelter. During this time, transportation became increasingly important leading to the development of the wheel and sailing. The construction of Pyramids in Egypt (circa 2700-2500 BC) might be considered the first instances of large structure constructions. Other ancient historic civil engineering constructions include the Parthenon by Iktinos in Ancient Greece (447-438 BC), the Appian Way by Roman engineers (c. 312 BC), and the Great Wall of China by General Meng T'ien under orders from Ch'in Emperor Shih Huang Ti (c. 220 BC). The Romans developed civil structures throughout their empire, including especially aqueducts, insulae, harbours, bridges, dams and roads.


The Archimedes' screw was operated by hand and could raise water efficiently.

Until modern times there was no clear distinction between civil engineering and architecture, and the term engineer and architect were mainly geographical variations referring to the same person, often used interchangeably. In the 18th century, the term civil engineering began to be used to and exchange, and in the construction of ports, harbours, moles, breakwaters and lighthouses, and in the art of distinguish it from military engineering.

The first self-proclaimed civil engineer was John Smeaton who constructed the Eddystone Lighthouse. In 1771 Smeaton and some of his colleagues formed the Smeatonian Society of Civil Engineers, a group of leaders of the profession who met informally over dinner. Though there was evidence of some technical meetings, it was little more than a social society.

In 1818 the Institution of Civil Engineers was founded in London, and in 1820 the eminent engineer Thomas Telford became its first president. The institution received a Royal Charter in 1828, formally recognising civil engineering as a profession. Its charter defined civil engineering as:

> "...the art of directing the great sources of power in nature for the use and convenience of man, as the means of production and of traffic in states, both for external and internal trade, as applied in the construction of roads, bridges, aqueducts, canals, river navigation and docks for internal intercourse navigation by artificial power for the purposes of commerce, and in the construction and application of machinery, and in the drainage of cities and towns."

The first degree in Civil Engineering in the United States was awarded by Rensselaer Polytechnic Institute in 1835.

History of the science of civil engineering


Pont du Gard, France, a Roman aqueduct built circa 19 BC.

Civil engineering is the application of physical and scientific principles, and its history is intricately linked to advances in understanding of physics and mathematics throughout history. Because civil engineering is a wide ranging profession, including several separate specialized sub-disciplines, its history is linked to knowledge of structures, materials science, geology, soils, hydrology, environment, mechanics and other fields.

Throughout ancient and medieval history most architectural design and construction was carried out by artisans, such as stone masons and carpenters, rising to the role of master builder. Knowledge was retained in guilds and seldom supplanted by advances. Structures, roads and infrastructure that existed were repetitive, and increases in scale were incremental.

One of the earliest examples of a scientific approach to physical and mathematical problems applicable to civil engineering is the work of Archimedes in the 3rd century BC, including Archimedes Principle, which underpins our understanding of buoyancy, and practical solutions such as Archimedes' screw.

The civil engineer

Education and licensure


The Institution of Civil Engineers headquarters in London

Civil engineers typically possess an academic degree with a major in civil engineering. The length of study for such a degree is usually four or five years and the completed degree is usually designated as a Bachelor of Engineering, though some universities designate the degree as a Bachelor of Science. The degree generally includes units covering physics, mathematics, project management, design and specific topics in civil engineering. Initially such topics cover most, if not all, of the sub-disciplines of civil engineering. Students then choose to specialize in one or more sub-disciplines towards the end of the degree.

Graduates can choose to pursue a postgraduate degree such as a Master of Engineering, Master of Science, or a Doctor of Philosophy in Engineering. The Master of Engineering degree may consist of either research, coursework or a mixture of the two. The Doctor of Philosophy consists of a significant research component and is often viewed as the entry point to academia. In the United Kingdom and various other European countries, the Master of Engineering is the minimum acceptable qualification for accreditation by the relevant professional bodies, and is often included as an extra year on the undergraduate engineering degree.

In most countries, a Bachelor's degree in engineering represents the first step towards professional certification and the degree program itself is certified by a professional body. After completing a certified degree program the engineer must satisfy a range of requirements (including work experience and exam requirements) before being certified. Once certified, the engineer is designated the title of Professional Engineer (in the United States, Canada and South Africa), Chartered Engineer (in most Commonwealth countries), Chartered Professional Engineer (in Australia and New Zealand), or European Engineer (in much of the European Union). There are international engineering agreements between relevant pressional bodies which are designed to allow engineers to practice across international borders.

The advantages of certification vary depending upon location. For example, in the United States and Canada "only a licensed engineer may prepare, sign and seal, and submit engineering plans and drawings to a public authority for approval, or seal engineering work for public and private clients.". This requirement is enforced by state and provincial legislation such as Quebec's Engineers Act. In other countries, no such legislation exists. In Australia, state licensing of engineers is limited to the state of Queensland. Practically all certifying bodies maintain a code of ethics that they expect all members to abide by or risk expulsion. In this way, these organizations play an important role in maintaining ethical standards for the profession. Even in jurisdictions where certification has little or no legal bearing on work, engineers are subject to contract law. In cases where an engineer's work fails he or she may be subject to the tort of negligence and, in extreme cases, the charge of criminal negligence. An engineer's work must also comply with numerous other rules and regulations such as building codes and legislation pertaining to environmental law.

Careers

There is no one typical career path for civil engineers. Most engineering graduates start with jobs of low responsibility, and as they prove their competence, they are given more and more responsible tasks, but within each subfield of civil engineering, and even within different segments of the market within each branch, the details of a career path can vary. In some fields and firms, entry-level engineers are put to work primarily monitoring construction in the field, serving as the "eyes and ears" of more senior design engineers; while in other areas, entry-level engineers end up performing the more routine tasks of analysis or design and interpretation. More senior engineers can move into doing more complex analysis or design work, or management of more complex design projects, or management of other engineers, or into specialized consulting, including forensic engineering.

Engineers are in high demand at banks, financial institutions and management consultancies because of their analytical skills.

Sub-disciplines

In general, civil engineering is concerned with the overall interface of human created fixed projects with the greater world. General civil engineers work closely with surveyors and specialized civil engineers to fit and serve fixed projects within their given site, community and terrain by designing grading, drainage, pavement, water supply, sewer service, electric and communications supply, and land divisions. General engineers spend much of their time visiting project sites, developing community consensus, and preparing construction plans. General civil engineering is also referred to as site engineering, a branch of civil engineering that primarily focuses on converting a tract of land from one usage to another. Civil engineers typically apply the principles of geotechnical engineering, structural engineering, environmental engineering, transportation engineering and construction engineering to residential, commercial, industrial and public works projects of all sizes and levels of construction.

Costal engineering

Coastal engineering is concerned with managing coastal areas.

Construction engineering

Building construction for several apartment blocks

Construction engineering involves planning and execution of the designs from transportation, site development, hydraulic, environmental, structural and geotechnical engineers. As construction firms tend to have higher business risk than other types of civil engineering firms, many construction engineers tend to take on a role that is more business-like in nature: drafting and reviewing contracts, evaluating logistical operations, and closely-monitoring prices of necessary supplies.

Environmental engineering

Environmental engineering deals with the treatment of chemical, biological, and/or thermal waste, the purification of water and air, and the remediation of contaminated sites, due to prior waste disposal or accidental contamination. Among the topics covered by environmental engineering are pollutant transport, water purification, waste water treatment, air pollution, solid waste treatment and hazardous waste management. Environmental engineers can be involved with pollution reduction, green engineering, and industrial ecology. Environmental engineering also deals with the gathering of information on the environmental consequences of proposed actions and the assessment of effects of proposed actions for the purpose of assisting society and policy makers in the decision making process.


A filter bed, a part of sewage treatment

Environmental engineering is the contemporary term for sanitary engineering, though sanitary engineering traditionally had not included much of the hazardous waste management and environmental remediation work covered by the term *environmental engineering*. Some other terms in use are public health engineering and environmental health engineering.

Geotechnical engineering

Geotechnical engineering is an area of civil engineering concerned with the rock and soil that civil engineering systems are supported by. Knowledge from the fields of geology, material science and testing, mechanics, and hydraulics are applied by geotechnical engineers to safely and economically design foundations, retaining walls, and similar structures. Environmental concerns in relation to groundwater and waste disposal have spawned a new area of study called geoenvironmental engineering where biology and chemistry are important.


A slab-on-grade foundation

Some of the unique difficulties of geotechnical engineering are the result of the variability and properties of soil. Boundary conditions are often well defined in other branches of civil engineering, but with soil, clearly defining these conditions can be impossible. The material properties and behaviour of soil are also difficult to predict due to the variability of soil and limited investigation. This contrasts with the relatively well defined material properties of steel and concrete used in other areas of civil engineering. Soil mechanics, which define the behaviour of soil, is complex due to stress-dependent material properties such as volume change, stress–strain relationship, and strength.

Materials engineering

Civil engineering also includes elements of materials engineering, also known as materials science. Construction materials with broad applications in civil engineering include ceramics such as Portland cement concrete (PCC) and hot mix asphalt concrete, metals such as aluminium and steel, and polymers such as polymethylmethacrylate (PMMA) and carbon fibers. Current research in these areas focus around increased strength, durability, workability, and reduced cost.

Structural engineering

Burj Dubai, the world's tallest building, currently under construction in Dubai

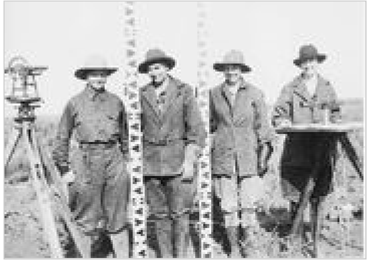
Structural engineering is concerned with the structural design and structural analysis of buildings, bridges, towers, flyovers, tunnels, off shore structures like oil and gas fields in the sea and other structures. This involves identifying the loads which act upon a structure and the forces and stresses which arise within that structure due to those loads, and then designing the structure to successfully support and resist those loads. The loads can be self weight of the structures, other dead load, live loads, moving (wheel) load, wind load, earthquake load, load from temperature change etc. The structural engineer must design structures to be safe for their users and to successfully fulfil the function they are designed for (to be *serviceable*). Due to the nature of some loading conditions, sub-disciplines within structural engineering have emerged, including wind engineering and earthquake engineering.

Clifton Suspension Bridge, designed by Isambard Kingdom Brunel, in Bristol, UK

Design considerations will include strength, stiffness and stability of the structure when subjected to loads which may be static, such as furniture or self-weight, or dynamic, such as wind, seismic, crowd or vehicle loads, or transitory, such as temporary construction loads or impact. Other considerations include cost, constructability, safety, aesthetics and sustainability.

Surveying

Surveying is the process by which a surveyor measures certain dimensions that generally occur on the surface of the Earth. Modern surveying equipment, such as electronic distance measurement (EDM), total stations, GPS surveying and laser scanning, allow for accurate measurement of angular deviation, horizontal, vertical and slope distances. This information is crucial to convert the data into a graphical representation of the Earth's surface, in the form of a map. This information is then used by civil engineers, contractors and even realtors to design from, build on, and trade, respectively. Elements of a building or structure must be correctly sized and positioned in relation to each other and to site boundaries and adjacent structures. Civil Engineers are trained in the basics of surveying and mapping, as well as geographic information systems.



An all-female surveying crew in Idaho, 1918

Transportation engineering

Transportation engineering is concerned with moving people and goods efficiently, safely, and in a manner conducive to a vibrant community. This involves specifying, designing, constructing, and maintaining transportation infrastructure which includes streets, canals, highways, rail systems, airports, ports, and mass transit. It includes areas such as transportation design, transportation planning, traffic engineering, urban engineering, queueing theory, pavement engineering, Intelligent Transportation System (ITS), and infrastructure management.

Clock

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Platform clock at King's Cross railway station, London.

A **clock** is an instrument for measuring, indicating and maintaining the time. The word *clock* is derived ultimately (via Dutch, Northern French, and Medieval Latin) from the Celtic words *clagan* and *clocca* meaning "bell". For horologists and other specialists the term *clock* continues to mean exclusively a device with a striking mechanism for announcing intervals of time acoustically, by ringing a bell, a set of chimes, or a gong. A silent instrument lacking such a mechanism has traditionally been known as a **timepiece**. In general usage today, however, a "clock" refers to any device for measuring and displaying the time which, unlike a watch, is not worn on the person.

History

The clock is one of the oldest human inventions, meeting the need to consistently measure intervals of time shorter than the natural units, the day, the lunar month, and the year. Such measurement requires devices. Devices operating on several different physical processes have been used over the millennia, culminating in the clocks of today.

Clock at the Royal Observatory, Greenwich

Sundials and other devices

The sundial, which measures the time of day by the direction of shadows cast by the sun, was widely used in ancient times. A well-designed sundial can measure local solar time with reasonable accuracy, and sundials continued to be used to monitor the performance of clocks until the modern era. However, its practical limitations - it requires the sun to shine and does not work at all during the night - encouraged the use of other techniques for measuring time.

Candle clocks and sticks of incense that burn down at, approximately, predictable speeds have also been used to estimate the passing of time. In an hourglass, fine sand pours through a tiny hole at a constant rate and indicates a predetermined passage of an arbitrary period of time.


Replica of an ancient Chinese incense clock

Water clocks

Water clocks, also known as clepsydrae(sg: clepsydra), along with the sundials, are possibly the oldest time-measuring instruments, with the only exceptions being the vertical gnomon and the day-counting tally stick. Given their great antiquity, where and when they first existed are not known and perhaps unknowable. The bowl-shaped outflow is the simplest form of a water clock and is known to have existed in Babylon and in Egypt around the 16th century BC. Other regions of the world, including India and China, also have early evidence of water clocks, but the earliest dates are less certain. Some authors, however, write about water clocks appearing as early as 4000 BC in these regions of the world.

A scale model of Su Song's Astronomical Clock Tower, built in 11th century Kaifeng, China. It was driven by a large waterwheel, chain drive, and escapement mechanism.

The Greek and Roman civilizations are credited for initially advancing water clock design to include complex gearing, which was connected to fanciful automata and also resulted in improved accuracy. These advances were passed on through Byzantium and Islamic times, eventually making their way to Europe. Independently, the Chinese developed their own advanced water clocks, passing their ideas on to Korea and Japan.

Some water clock designs were developed independently and some knowledge was transferred through the spread of trade. It is important to point out that the need for the common person to 'know what time it is' largely did not exist until the Industrial Revolution, when it became important to keep track of hours worked. In the earliest of times, however, the purpose for using a water clock was for astronomical and astrological reasons. These early water clocks were calibrated with a sundial. Through the centuries, water clocks were used for timing lawyer's speeches during a trial, labors of prostitutes, night watches of guards, sermons and Masses in church, to name only a few. While never reaching the level of accuracy based on today's standards of timekeeping, the water clock was the most accurate and commonly used timekeeping device for millennia, until it was replaced by the more accurate pendulum clock in 17th century Europe.

Early clocks

In 797 (or possibly 801), the Abbasid caliph of Baghdad, Harun al-Rashid, presented Janae with an Asian Elephant named Abul-Abbas together with a "particularly elaborate example" of a water clock.

None of the first clocks survived from 13th century Europe, but various mentions in church records reveal some of the early history of the clock.

Medieval religious institutions required clocks to measure and indicate the passing of time because, for many centuries, daily prayer and work schedules had to be strictly regulated. This was done by various types of time-telling and recording devices, such as water clocks, sundials and marked candles, probably used in combination. Important times and durations were broadcast by bells, rung either by hand or by some mechanical device such as a falling weight or rotating beater.

The word *horologia* (from the Greek ὥρα, hour, and λεγειν, to tell) was used to describe all these devices, but the use of this word (still used in several romance languages) for all timekeepers conceals from us the true nature of the mechanisms. For example, there is a record that in 1176 Sens Cathedral installed a ‘horologe’ but the mechanism used is unknown. According to Jocelin of Brakelond, in 1198 during a fire at the abbey of St Edmundsbury (now Bury St Edmunds), the monks 'ran to the clock' to fetch water, indicating that their water clock had a reservoir large enough to help extinguish the occasional fire .

These early clocks may not have used hands or dials, but “told” the time with audible signals.

A new mechanism

The word *clock* (from the Latin word *clocca*, "bell"), which gradually supersedes "horologe", suggests that it was the sound of bells which also characterized the prototype mechanical clocks that appeared during the 13th century in Europe.

Between 1280 and 1320, there is an increase in the number of references to clocks and horologes in church records, and this probably indicates that a new type of clock mechanism had been devised. Existing clock mechanisms that used water power were being adapted to take their driving power from falling weights. This power was controlled by some form of oscillating mechanism, probably derived from existing bell-ringing or alarm devices. This controlled release of power - the escapement - marks the beginning of the true mechanical clock.

Outside of Europe, the escapement mechanism had been known and used in medieval China, as the Song Dynasty horologist and engineer Su Song (1020 - 1101) incorporated it into his astronomical clock-tower of Kaifeng in 1088. However, his astronomical clock and rotating armillary sphere still relied on the use of flowing water (ie. hydraulics), while European clockworks of the following centuries shed this old habit for a more efficient driving power of weights, in addition to the escapement mechanism.

In the 13th century, clock construction and engineering entered a new phase with the advancements made by Al-Jazari, a Muslim engineer from Diyar-Bakr in South East Turkey, who is thought to be behind the birth to the concept of automatic machines. While working for Artuqid king of Diyar-Bakr, Nasir al-Din, al-Jazari made numerous clocks of all shapes and sizes. In 1206 he was ordered by the king to document his inventions leading to the publication of an outstanding book on engineering called "The Book of Knowledge of Ingenious Mechanical Devices”. This book became an invaluable resource for people of different engineering backgrounds as it described 50 mechanical devices in 6 categories, including water clocks. The most reputed clocks included the Elephant, the Castle and Scribe clocks, all of which were reconstructed by Muslim Heritage Consulting for Ibn Battuta Shopping Mall in Dubai (UAE), where they are fully functional. As well as telling the time, these grand clocks were symbols of status, grandeur and wealth of the Urtuq State.

These mechanical clocks were intended for two main purposes: for signalling and notification (e.g. the timing of services and public events), and for modeling the solar system. The former purpose is administrative, the latter arises naturally given the scholarly interest in astronomy, science, astrology, and how these subjects integrated with the religious philosophy of the time. The astrolabe was used both by astronomers and astrologers, and it was natural to apply a clockwork drive to the rotating plate to produce a working model of the solar system.

Simple clocks intended mainly for notification were installed in towers, and did not always require dials or hands. They would have announced the canonical hours or intervals between set times of prayer. Canonical hours varied in length as the times of sunrise and sunset shifted. The more sophisticated astronomical clocks would have had moving dials or hands, and would have shown the time in various time systems, including Italian hours, canonical hours, and time as measured by astronomers at the time. Both styles of clock started acquiring extravagant features such as automata.

In 1283, a large clock was installed at Dunstable Priory; its location above the rood screen suggests that it was not a water clock. In 1292, Canterbury Cathedral installed a 'great horloge'. Over the next 30 years there are brief mentions of clocks at a number of ecclesiastical institutions in England, Italy, and France. In

1322, a new clock was installed in Norwich, an expensive replacement for an earlier clock installed in 1273. This had a large (2 metre) astronomical dial with automata and bells. The costs of the installation included the full-time employment of two clockkeepers for two years.

Early astronomical clocks


Richard of Wallingford pointing to a clock, his gift to St Albans Abbey

Besides the Chinese astronomical clock of Su Song in 1088 mentioned above, in Europe there were the clocks constructed by Richard of Wallingford in St Albans by 1336, and by Giovanni de Dondi in Padua from 1348 to 1364. They no longer exist, but detailed descriptions of their design and construction survive, while modern reproductions have been made. They illustrate how quickly the theory of the mechanical clock had been translated into practical constructions, and also that one of the many impulses to their development had been the desire of astronomers to investigate celestial phenomena.

Wallingford's clock had a large astrolabe-type dial, showing the sun, the moon's age, phase, and node, a star map, and possibly the planets. In addition, it had a wheel of fortune and an indicator of the state of the tide at London Bridge. Bells rang every hour, the number of strokes indicating the time.

Dondi's clock was a seven-sided construction, 1 metre high, with dials showing the time of day, including minutes, the motions of all the known planets, an automatic calendar of fixed and movable feasts, and an eclipse prediction hand rotating once every 18 years.

It is not known how accurate or reliable these clocks would have been. They were probably adjusted manually every day to compensate for errors caused by wear and imprecise manufacture.

The Salisbury Cathedral clock, built in 1386, is considered to be the world's oldest surviving mechanical clock that strikes the hours.

Later developments

Clockmakers developed their art in various ways. Building smaller clocks was a technical challenge, as was improving accuracy and reliability. Clocks could be impressive showpieces to demonstrate skilled craftsmanship, or less expensive, mass-produced items for domestic use. The escapement in particular was an important factor affecting the clock's accuracy, so many different mechanisms were tried.

Spring-driven clocks appeared during the 1400s, although they are often erroneously credited to Nürnberg watchmaker Peter Henlein (or Henle, or Hele) around 1511. The earliest existing spring driven clock is the chamber clock given to Peter the Good, Duke of Burgundy, around 1430, now in the Germanisches Nationalmuseum. Spring power presented clockmakers with a new problem; how to keep the clock movement running at a constant rate as the spring ran down. This resulted in the invention of the *stackfreed* and the fusee in the 1400s, and many other innovations, down to the invention of the modern *going barrel* in 1760.

The first record of a minute hand on a clock is 1475, in the Almanus Manuscript of Brother Paul.

During the 15th and 16th centuries, clockmaking flourished, particularly in the metalworking towns of Nuremberg and Augsburg, and in France, Blois. Some of the more basic table clocks have only one time-keeping hand, with the dial between the hour markers being divided into four equal parts making the clocks readable to the nearest 15 minutes. Other clocks were exhibitions of craftsmanship and skill, incorporating astronomical indicators and musical movements. The cross-beat escapement was developed in 1585 by Jost Burgi, who also developed the remontoire. Burgi's accurate clocks helped Tycho Brahe to observe astronomical events with much greater precision than before.

The first record of a second hand on a clock is about 1560, on a clock now in the Fremersdorf collection. However, this clock could not have been accurate, and the second hand was probably for indicating that the clock was working.


French rococo bracket clocks, (Museum of Time, Besançon)

The next development in accuracy occurred after 1657 with the invention of the pendulum clock. Galileo had the idea to use a swinging bob to regulate the motion of a time telling device earlier in the 17th century. Christiaan Huygens, however, is usually credited as the inventor. He determined the mathematical formula that related pendulum length to time (99.38 cm or 39.13 inches for the one second movement) and had the first pendulum-driven clock made. In 1670, the English clockmaker William Clement created the anchor escapement, an improvement over Huygens' crown escapement. Within just one generation, minute hands and then second hands were added.

A major stimulus to improving the accuracy and reliability of clocks was the importance of precise time-keeping for navigation. The position of a ship at sea could be determined with reasonable accuracy if a navigator could refer to a clock that lost or gained less than about 10 seconds per day. This clock could not contain a pendulum, which would be virtually useless on a rocking ship. Many European governments offered a large prize for anyone that could determine longitude accurately; for example, Great Britain offered 20,000 pounds, equivalent to millions of dollars today. The reward was eventually claimed in 1761 by John Harrison, who dedicated his life to improving the accuracy of his clocks. His H5 clock is reported to have lost less than 5 seconds over 10 days.

The excitement over the pendulum clock had attracted the attention of designers resulting in a proliferation of clock forms. Notably, the longcase clock (also known as the *grandfather clock*) was created to house the pendulum and works. The English clockmaker William Clement is also credited with developing this form in 1670 or 1671. It was also at this time that clock cases began to be made of wood and clock faces to utilize enamel as well as hand-painted ceramics.

On November 17, 1797, Eli Terry received his first patent for a clock. Terry is known as the founder of the American clock-making industry.

French decimal clock from the time of the French Revolution

Alexander Bain, Scottish clockmaker, patented the electric clock in 1840. The electric clock's mainspring is wound either with an electric motor or with an electro-magnet and armature. In 1841, he first patented the electromagnetic pendulum.

The development of electronics in the twentieth century led to clocks with no clockwork parts at all. Time in these cases is measured in several ways, such as by the vibration of a tuning fork, the behaviour of quartz crystals, the resonance of polycarbonates., or the quantum vibrations of atoms. Even mechanical clocks have since come to be largely powered by batteries, removing the need for winding.

How clocks work

The invention of the mechanical clock in the 13th century started a change in timekeeping methods from continuous processes, such as the motion of the gnomon's shadow on a sundial or the flow of liquid in a water clock, to repetitive oscillatory processes, like the swing of a pendulum or the vibration of a quartz crystal, which were more accurate. All modern clocks use oscillation.

Although the methods they use vary, all oscillating clocks, mechanical and digital and atomic, work similarly and can be divided into analogous parts. They consist of an object that repeats the same motion over and over again, an *oscillator*, with a precisely constant time interval between each repetition, or 'beat'. Attached to the oscillator is a *controller* device, which sustains the oscillator's motion by replacing the energy it loses to friction, and converts its oscillations into a series of pulses. The pulses are then added up in a chain of some type of *counter*s to express the time in convenient units, usually seconds, minutes, hours, etc. Then finally some kind of *indicator* displays the result in a human-readable form.

Oscillator

The timekeeping element in every modern clock is a harmonic oscillator, a physical object (resonator) that vibrates or oscillates repetitively at a precisely constant frequency.

- In mechanical clocks, this is either a pendulum or a balance wheel.
- In some early electronic clocks and watches such as the Accutron, it is a tuning fork.
- In quartz clocks and watches, it is a quartz crystal.
- In atomic clocks, it is the vibration of electrons in atoms as they emit microwaves.
- In early mechanical clocks before 1657, it was a crude balance wheel or foliot which was not a harmonic oscillator because it lacked a balance spring. As a result they were very inaccurate, with errors of perhaps an hour a day.

The advantage of a harmonic oscillator over other forms of oscillator is that it employs resonance to vibrate at a precise natural resonant frequency or 'beat' dependent only on its physical characteristics, and resists vibrating at other rates. The possible precision achievable by a harmonic oscillator is measured by a parameter called its Q, or quality factor, which increases (other things being equal) with its resonant frequency. This is why there has been a long term trend toward higher frequency oscillators in clocks.

Some clocks rely for their accuracy on an external oscillator; that is, they are automatically synchronized to a more accurate clock:

- Slave clocks, used in large institutions and schools from the 1860s to the 1970s, kept time with a pendulum, but were wired to a master clock in the building, and periodically received a signal to synchronize them with the master, often on the hour.
- Synchronous electric clocks don't have an internal oscillator, but rely on the 50 or 60 Hz oscillation of the AC power line, which is synchronized by the utility to a precision oscillator. This drives a synchronous motor in the clock which rotates once for every cycle of the line voltage, and drives the gear train.
- Computer real time clocks keep time with a quartz crystal, but are periodically (usually weekly) synchronized over the internet to atomic clocks (UTC), using a system called Network Time Protocol.
- Radio clocks keep time with a quartz crystal, but are periodically (often daily) synchronized to atomic clocks (UTC) with time signals from government radio stations like WWV, WWVB, CHU, DCF77 and the GPS system.

Controller

This has the dual function of keeping the oscillator running by giving it 'pushes' to replace the energy lost to friction, and converting its vibrations into a series of pulses that serve to measure the time.

- In mechanical clocks, this is the escapement, which gives precise pushes to the swinging pendulum or balance wheel, and releases one gear tooth of the *escape wheel* at each swing, allowing all the clocks wheels to move forward a fixed amount with each swing.
- In electronic clocks this is an electronic oscillator circuit that gives the vibrating quartz crystal or tuning fork tiny 'pushes', and generates a series of electrical pulses, one for each vibration of the oscillator, which is called the clock signal.
- In atomic clocks the controller is an evacuated microwave cavity attached to a microwave oscillator controlled by a microprocessor. A thin gas of cesium atoms is released into the cavity where they are exposed to microwaves. A laser measures how many atoms have absorbed the microwaves, and an electronic feedback control system called a phase locked loop tunes the microwave oscillator until it is at the exact frequency that causes the atoms to vibrate and absorb the microwaves. Then the microwave signal is divided by digital counters to become the clock signal.

In mechanical clocks, the low Q of the balance wheel or pendulum oscillator made them very sensitive to the disturbing effect of the impulses of the escapement, so the escapement had a great effect on the accuracy of the clock, and many escapement designs were tried. The higher Q of resonators in electronic clocks makes them relatively insensitive to the disturbing effects of the drive power, so the driving oscillator circuit is a much less critical component.

Counter chain

This counts the pulses and adds them up to get traditional time units of seconds, minutes, hours, etc. It usually has a provision for *setting* the clock by manually entering the correct time into the counter.

- In mechanical clocks this is done analogically by a gear train, also called wheel train. The gear train also has a second function; to transmit mechanical power from the power source to run the oscillator. There is a friction coupling called the 'cannon pinion' between the gears driving the hands and the rest of the clock, allowing the hands to be turned by a knob on the back to set the time.
- In digital clocks a series of integrated circuit counters or dividers add the pulses up digitally, using binary logic. Often pushbuttons on the case allow the hour and minute counters to be incremented and decremented to set the time.

Indicator

This displays the count of seconds, minutes, hours, etc. in a human readable form.

- The earliest mechanical clocks in the 13th century didn't have a visual indicator and signalled the time audibly by striking bells. Many clocks to this day are striking clocks which chime the hours.
- Analog clocks, including almost all mechanical and some electronic clocks, have a traditional dial or clock face, that displays the time in analog form with moving hour and minute hand.
- Digital clocks display the time in periodically changing digits on a digital display.
- Talking clocks and the speaking clock services provided by telephone companies speak the time audibly, using either recorded or digitally synthesized voices.

Types

Clocks can be classified by the type of time display, as well as by the method of timekeeping.

Time display methods

Analogue clocks

Analogue clocks usually indicate time using angles. The most common clock face uses a fixed numbered dial or dials and moving hand or hands. It usually has a circular scale of 12 hours, which can also serve as a scale of 60 minutes, and 60 seconds if the clock has a second hand. Many other styles and designs have been used throughout the years, including dials divided into 6, 8, 10, and 24 hours. The only other widely used clock face today is the 24 hour analogue dial, because of the use of 24 hour time in military organizations and timetables. The 10-hour clock was briefly popular during the French Revolution, when the metric system was applied to time measurement, and an Italian 6 hour clock was developed in the 18th century, presumably to save power (a clock or watch chiming 24 times uses more power).

A linear clock at London's Piccadilly Circus tube station. The 24 hour band moves across the static map, keeping pace with the apparent movement of the sun above ground, and a pointer fixed on London points to the current time

Another type of analogue clock is the sundial, which tracks the sun continuously, registering the time by the shadow position of its gnomon. Sundials use some or part of the 24 hour analogue dial. There also exist clocks which use a digital display despite having an analogue mechanism—these are commonly referred to as flip clocks.

Alternative systems have been proposed. For example, the TWELV clock indicates the current hour using one of twelve colors, and indicates the minute by showing a proportion of a circular disk, similar to a moon phase.

The mechanics of analogue clocks were also the subject of the Grammy Award winning Coldplay single, *Clocks* in which the continual ticking of the clocks mesmerises and fascinates the narrator of the song.

Digital clocks

Digital clocks display a numeric representation of time. Two numeric display formats are commonly used on digital clocks:

- the 24-hour notation with hours ranging 00–23;
- the 12-hour notation with AM/PM indicator, with hours indicated as 12AM, followed by 1AM–11AM, followed by 12PM, followed by 1PM–11PM (a notation mostly used in the United States).

Digital clock outside Kanazawa Station displaying the time by controlling valves on a fountain

Most digital clocks use an LCD, LED, or VFD display; many other display technologies are used as well (cathode ray tubes, nixie tubes, etc.). After a reset, battery change or power failure, digital clocks without a backup battery or capacitor either start counting from 00:00, or stay at 00:00, often with blinking digits indicating that time needs to be set. Some newer clocks will actually reset themselves based on radio or Internet time servers that are tuned to national atomic clocks. Since the release of digital clocks in the mainstream, the use of analogue clocks has dropped dramatically.

Auditory clocks

Basic digital clock radio

For convenience, distance, telephony or blindness, auditory clocks present the time as sounds. The sound is either spoken natural language, (e.g. "The time is twelve thirty-five"), or as auditory codes (e.g. number of sequential bell rings on the hour represents the number of the hour like the clock Big Ben). Most telecommunication companies also provide a Speaking clock service as well.

Purposes

Clocks are in homes, offices and many other places; smaller ones (watches) are carried on the wrist; larger ones are in public places, e.g. a train station or church. A small clock is often shown in a corner of computer displays, mobile phones and many MP3 players.

The purpose of a clock is not always to *display* the time. It may also be used to *control* a device according to time, e.g. an alarm clock, a VCR, or a time bomb (see: counter). However, in this context, it is more appropriate to refer to it as a timer or trigger mechanism rather than strictly as a clock.

Computers depend on an accurate internal clock signal to allow synchronized processing. (A few research projects are developing CPUs based on asynchronous circuits.) Some computers also maintain time and date for all manner of operations whether these be for alarms, event initiation, or just to display the time of day. The internal computer clock is generally kept running by a small battery. Many computers will still function even if the internal clock battery is dead, but the computer clock will need to be reset each time the computer is restarted, since once power is lost, time is also lost.

Ideal clocks

An ideal clock is a scientific principle that measures the ratio of the duration of natural processes, and thus will give the time measure for use in physical theories. Therefore, to define an ideal clock in terms of any physical theory would be circular. An ideal clock is more appropriately defined in relationship to the set of all physical processes. An ideal clock should too measure time in consistent, for example decimalized time units.

This leads to the following definitions:

- A clock is a recurrent process and a counter.
- A good clock is one which, when used to measure other recurrent processes, finds many of them to be periodic.
- An ideal clock is a clock (i.e., recurrent process) that makes the most other recurrent processes periodic.

The recurrent, periodic process (e.g. a metronome) is an oscillator and typically generates a *clock signal*. Sometimes that signal alone is (confusingly) called "the clock", but sometimes "the clock" includes the counter, its indicator, and everything else supporting it.

This definition can be further improved by the consideration of successive levels of smaller and smaller error tolerances. While not all physical processes can be surveyed, the definition should be based on the set of physical processes which includes all individual physical processes which are proposed for consideration. Since atoms are so numerous and since, within current measurement tolerances they all beat in a manner such that if one is chosen as periodic then the others are all deemed to be periodic also, it follows that atomic clocks represent ideal clocks to within present measurement tolerances and in relation to all presently known physical processes. However, they are not so designated by fiat. Rather, they are designated as the current ideal clock because they are currently the best instantiation of the definition.

Navigation


John Harrison's Chronometer H5

Navigation by ships depends on the ability to measure latitude and longitude. Latitude is fairly easy to determine through celestial navigation, but the measurement of longitude requires accurate measurement of time. This need was a major motivation for the development of accurate mechanical clocks. John Harrison created the first highly accurate marine chronometer in the mid-18th century. The Noon gun in Cape Town still fires an accurate signal to allow ships to check their chronometers.

Use of a common clock in radio signal producing satellites is fundamental to the operation of GPS (Global Positioning System) navigation devices.

Seismology

In determining the location of an earthquake, the arrival time of several types of seismic wave at at least four dispersed observers is dependent upon each observer recording wave arrival times according to a common clock.

Specific types of clocks

- Alarm clock
- Flip clock
- Astronomical clock
- Atomic clock
- Balloon clock
- Binary clock
- Bracket clock
- Carriage clock
- Cartel clock
- Chiming clock
- Clock network
- Floral clock
- Game clock
- Hourglass
- Japanese clock
- Lantern clock
- Lighthouse Clock
- Longcase (or "grandfather") clock
- Master clock
- Mantel clock
- Musical clock
- Paper clock
- Sidereal clock
- Skeleton clock
- Slave clock
- Speaking clock
- Stopwatch
- Striking clock
- Sundial
- Talking clock
- Tall-case clock
- Tide clock
- Time ball

- Clock of the Long Now
- Clock tower
- Countdown clock
- Cuckoo clock
- Data clock for timescapes created with time-technology
- Digital clock
- Doll's head clock
- Electric clock
- Pedestal clock
- Pendulum clock
- Projection clock
- Quartz clock
- Radio clock
- Railroad chronometer
- Reference clock
- Rolling ball clock
- Time clock
- Torsion pendulum clock
- Tower clock
- Wall clock
- Watch
- Water clock
- World clock

Coal

2008/9 Schools Wikipedia Selection. Related subjects: Environment; Mineralogy

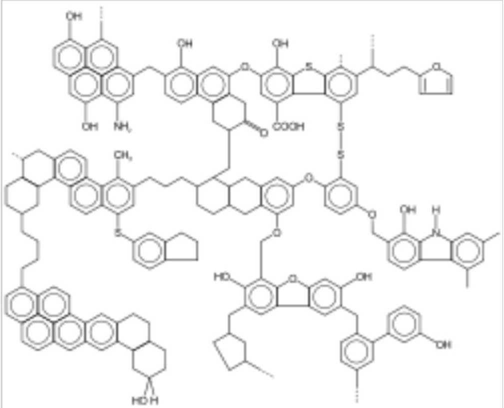
Coal is a fossil fuel formed in ecosystems where plant remains were preserved by water and mud from oxidization and biodegradation, thus sequestering atmospheric carbon. Coal is a readily combustible black or brownish-black rock. It is a sedimentary rock, but the harder forms, such as anthracite coal, can be regarded as metamorphic rocks because of later exposure to elevated temperature and pressure. It is composed primarily of carbon and hydrogen along with small quantities of other elements, notably sulfur. It is the largest source of fuel for generation of electricity world-wide, as well as the largest world-wide source of carbon dioxide emissions, which according to the IPCC, contribute to climate change and global warming. In terms of carbon dioxide emissions, coal is slightly ahead of petroleum and about double that of natural gas. Coal is extracted from the ground by coal mining, either underground mining or open pit mining (surface mining).


Coal

Types of coal

As geological processes apply pressure to dead biotic matter over time, under suitable conditions it is transformed successively into

- Peat, considered to be a precursor of coal. It has industrial importance as a fuel in some countries, for example, Ireland and Finland.
- Lignite, also referred to as brown coal, is the lowest rank of coal and used almost exclusively as fuel for electric power generation. Jet is a compact form of lignite that is sometimes polished and has been used as an ornamental stone since the Iron Age.
- Sub-bituminous coal, whose properties range from those of lignite to those of bituminous coal and are used primarily as fuel for steam-electric power generation. Additionally, it is an important source of light aromatic hydrocarbons for the chemical synthesis industry.
- Bituminous coal, a dense mineral, black but sometimes dark brown, often with well-defined bands of bright and dull material, used primarily as fuel in steam-electric power generation, with substantial quantities also used for heat and power applications in manufacturing and to make coke.
- Anthracite, the highest rank; a harder, glossy, black coal used primarily for residential and commercial space heating. It may be divided further in to metamorphically altered bituminous coal and *petrified oil*, as from the deposits in Pennsylvania.


Example chemical structure of coal

- Graphite, technically the highest rank, but difficult to ignite and is not so commonly used as fuel: it is mostly used in pencils and, when powdered, as a lubricant.

The classification of coal is generally based on the content of volatiles. However, the exact classification varies between countries. According to the German classification, coal is classified as follows:

Name	Volatiles %	C Carbon %	H Hydrogen %	O Oxygen %	S Sulfur %	Heat content kJ/kg
Braunkohle (Lignite)	45-65	60-75	6.0-5.8	34-17	0.5-3	<28470
Flammkohle (Flame coal)	40-45	75-82	6.0-5.8	>9.8	~1	<32870
Gasflammkohle (Gas flame coal)	35-40	82-85	5.8-5.6	9.8-7.3	~1	<33910
Gaskohle (Gas coal)	28-35	85-87.5	5.6-5.0	7.3-4.5	~1	<34960
Fettkohle (Fat coal)	19-28	87.5-89.5	5.0-4.5	4.5-3.2	~1	<35380
Esskohle (Forge coal)	14-19	89.5-90.5	4.5-4.0	3.2-2.8	~1	<35380
Magerkohle (Non baking coal)	10-14	90.5-91.5	4.0-3.75	2.8-3.5	~1	35380
Anthrazit (Anthracite)	7-12	>91.5	<3.75	<2.5	~1	<35300

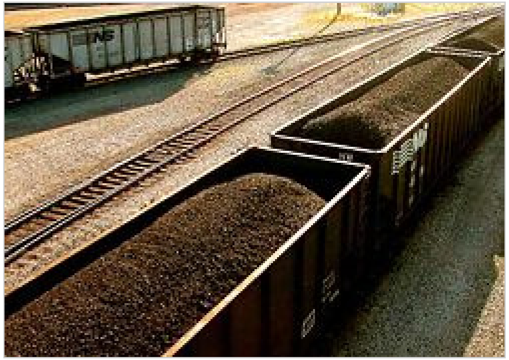
The middle six grades in the table represent a progressive transition from the English-language sub-bituminous to bituminous coal, while the last class is an approximate equivalent to anthracite, but more inclusive (the U.S. anthracite has < 8% volatiles).

Early use

China Coal Information Institute reports the Chinese mined coalstone for fuel 10,000 years ago at the time of the New Stone Age, or Neolithic Era. "People in Shanxi, now the largest coal production base, have been burning coal as fuel since then." Outcrop coal was used in Britain during the Bronze Age (2000-3000 years BC), where it has been detected as forming part of the composition of funeral pyres. It was also commonly used in the early period of the Roman occupation: Evidence of trade in coal (dated to about AD 200) has been found at the inland port of Heronbridge, near Chester, and in the Fenlands of East Anglia, where coal from the Midlands was transported via the Car Dyke for use in drying grain. Coal cinders have been found in the hearths of villas and military forts, particularly in Northumberland, dated to around AD 400. In the west of England contemporary writers described the wonder of a permanent brazier of coal on the altar of Minerva at *Aquae Sulis* (modern day Bath) although in fact easily-accessible surface coal from what is now the Somerset coalfield was in common use in quite lowly dwellings locally.

However, there is no evidence that the product was of great importance in Britain before the High Middle Ages, after about AD 1000. Mineral coal came to be referred to as "seacoal," probably because it came to many places in eastern England, including London, by sea. This is accepted as the more likely explanation for the name than that it was found on beaches, having fallen from the exposed coal seams above or washed out of underwater coal seam outcrops. These easily accessible sources had largely become exhausted (or could not meet the growing demand) by the 13th century, when underground mining from shafts or adits was developed. In London there is still a Seacoal Lane (off the north side of Ludgate Hill) where the coal merchants used to conduct their business. An alternative name was "pitcoal," because it came from mines. It was, however, the development of the Industrial Revolution that led to the large-scale use of coal, as the steam engine took over from the water wheel.

Uses today



Coal rail cars in Ashtabula, Ohio.

Coal as fuel

Coal is primarily used as a solid fuel to produce electricity and heat through combustion. World coal consumption is about 6.2 billion tons annually, of which about 75% is used for the production of electricity. China produced 2.38 billion tons in 2006 and India produced about 447.3 million tons in 2006. 68.7% of China's electricity comes from coal. The USA consumes about 1.053 billion tons of coal each year, using 90% of it for generation of electricity. The world in total produced 6.19 billion tons of coal in 2006.

When coal is used for electricity generation, it is usually pulverized and then burned in a furnace with a boiler. The furnace heat converts boiler water to steam, which is then used to spin turbines which turn generators and create electricity. The thermodynamic efficiency of this process has been improved over time. "Standard" steam turbines have topped out with some of the most advanced reaching about 35% thermodynamic efficiency for the entire process, which means 65% of the coal energy is waste heat released into the surrounding environment. Old coal power plants, especially "grandfathered" plants, are significantly less efficient and produce higher levels of waste heat.

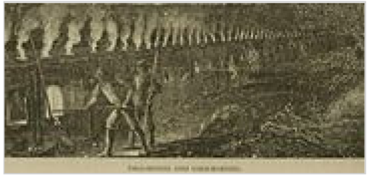
The emergence of the supercritical turbine concept envisions running a boiler at extremely high temperatures and pressures with projected efficiencies of 46%, with further theorized increases in temperature and pressure perhaps resulting in even higher efficiencies.

Other efficient ways to use coal are combined cycle power plants, combined heat and power cogeneration, and an MHD topping cycle.

Approximately 40% of the world electricity production uses coal. The total known deposits recoverable by current technologies, including highly polluting, low energy content types of coal (i.e., lignite, bituminous), might be sufficient for 300 years' use at current consumption levels, although maximal production could be reached within decades (see World Coal Reserves, below).

A more energy-efficient way of using coal for electricity production would be via solid-oxide fuel cells or molten-carbonate fuel cells (or any oxygen ion transport based fuel cells that do not discriminate between fuels, as long as they consume oxygen), which would be able to get 60%–85% combined efficiency (direct electricity + waste heat steam turbine). Currently these fuel cell technologies can only process gaseous fuels, and they are also sensitive to sulfur poisoning, issues which would first have to be worked out before large scale commercial success is possible with coal. As far as gaseous fuels go, one idea is pulverized coal in a gas carrier, such as nitrogen. Another option is coal gasification with water, which may lower fuel cell voltage by introducing oxygen to the fuel side of the electrolyte, but may also greatly simplify carbon sequestration.

Coking and use of coke



Coke burning

Coke is a solid carbonaceous residue derived from low-ash, low-sulfur bituminous coal from which the volatile constituents are driven off by baking in an oven without oxygen at temperatures as high as 1,000 °C (1,832 °F) so that the fixed carbon and residual ash are fused together. Metallurgic coke is used as a fuel and as a reducing agent in smelting iron ore in a blast furnace. Coke from coal is grey, hard, and porous and has a heating value of 24.8 million Btu/ton (29.6 MJ/kg). Some cokemaking processes produce valuable by-products that include coal tar, ammonia, light oils, and " coal gas".

Petroleum coke is the solid residue obtained in oil refining, which resembles coke but contains too many impurities to be useful in metallurgical applications.

Gasification

High prices of oil and natural gas are leading to increased interest in "BTU Conversion" technologies such as gasification, methanation and liquefaction.

Coal gasification breaks down the coal into smaller molecular weight molecules, usually by subjecting it to high temperature and pressure, using steam and measured amounts of oxygen. This leads to the production of syngas, a mixture mainly consisting of carbon monoxide (CO) and hydrogen (H_2).

In the past, coal was converted to make coal gas, which was piped to customers to burn for illumination, heating, and cooking. At present, the safer natural gas is used instead. South Africa still uses gasification of coal for much of its petrochemical needs.

The Synthetic Fuels Corporation was a U.S. government-funded corporation established in 1980 to create a market for alternatives to imported fossil fuels (such as coal gasification). The corporation was discontinued in 1985.

Gasification is also a possibility for future energy use, as the produced syngas can be cleaned-up relatively easily leading to cleaner burning than burning coal directly (the conventional way). The cleanliness of the cleaned-up syngas is comparable to natural gas enabling to burn it in a more efficient gas turbine rather than in a boiler used to drive a steam turbine. Syngas produced by gasification can be CO-shifted meaning that the combustible CO in the syngas is transferred into carbon dioxide (CO_2) using water as a reactant. The CO-shift reaction also produces an amount of combustible hydrogen (H_2) equal to the amount of CO converted into CO_2. The CO_2 concentrations (or rather CO_2 partial pressures) obtained by using coal gasification followed by a CO-shift reaction are much higher than in case of direct combustion of coal in air (which is mostly nitrogen). These higher concentrations of carbon dioxide make carbon capture and storage much more economical than it otherwise would be.

Liquefaction - Coal-To-Liquids (CTL)

Coals can also be converted into liquid fuels like gasoline or diesel by several different processes. The Fischer-Tropsch process of indirect synthesis of liquid hydrocarbons was used in Nazi Germany for many years and is today used by Sasol in South Africa. Coal would be gasified to make syngas (a balanced purified mixture of CO and H_2 gas) and the syngas condensed using Fischer-Tropsch catalysts to make light hydrocarbons which are further processed into gasoline and diesel. Syngas can also be converted to methanol, which can be used as a fuel, fuel additive, or further processed into gasoline via the Mobil M-gas process.

A direct liquefaction process Bergius process (liquefaction by hydrogenation) is also available but has not been used outside Germany, where such processes were operated both during World War I and World War II. SASOL in South Africa has experimented with direct hydrogenation. Several other direct liquefaction processes have been developed, among these being the SRC-I and SRC-II (Solvent Refined Coal) processes developed by Gulf Oil and implemented as pilot plants in the United States in the 1960s and 1970s.

Another direct hydrogenation process was explored by the NUS Corporation in 1976 and patented by Wilburn C. Schroeder. The process involved dried, pulverized coal mixed with roughly 1wt% molybdenum catalysts. Hydrogenation occurred by use of high temperature and pressure synthesis gas produced in a separate gasifier. The process ultimately yielded a synthetic crude product, Naphtha, a limited amount of C_3/C_4 gas, light-medium weight liquids (C_5-C_{10}) suitable for use as fuels, small amounts of NH_3 and significant amounts of CO_2.

Yet another process to manufacture liquid hydrocarbons from coal is low temperature carbonization (LTC). Coal is coked at temperatures between 450 and 700°C compared to 800 to 1000°C for metallurgical coke. These temperatures optimize the production of coal tars richer in lighter hydrocarbons than normal coal tar. The coal tar is then further processed into fuels. The Karrick process was developed by Lewis C. Karrick, an oil shale technologist at the U.S. Bureau of Mines in the 1920s.

All of these liquid fuel production methods release carbon dioxide (CO_2) in the conversion process, far more than is released in the extraction and refinement of liquid fuel production from petroleum. If these methods were adopted to replace declining petroleum supplies, carbon dioxide emissions would be greatly increased on a global scale. For future liquefaction projects, Carbon dioxide sequestration is proposed to avoid releasing it into the atmosphere, though no pilot projects have confirmed the feasibility of this approach on a wide scale. As CO_2 is one of the process streams, sequestration is easier than from flue gases produced in combustion of coal with air, where CO_2 is diluted by nitrogen and other gases. Sequestration will, however, add to the cost.

The reaction of coal and water using high temperature heat from a nuclear reactor offers promise of liquid transport fuels that could prove carbon-neutral compared to petroleum use. The development of a reliable nuclear reactor that could provide 900 to 1000 deg C process heat, such as the pebble bed reactor, would be necessary.

Coal liquefaction is one of the backstop technologies that could potentially limit escalation of oil prices and mitigate the effects of transportation energy shortage that some authors have suggested could occur under peak oil. This is contingent on liquefaction production capacity becoming large enough to satiate the very large and growing demand for petroleum. Estimates of the cost of producing liquid fuels from coal suggest that domestic U.S. production of fuel from coal becomes cost-competitive with oil priced at around 35 USD per barrel, (break-even cost). The current price of oil, as of July 11, 2008, is 145 USD per barrel. This makes coal a viable financial alternative to oil for the time being, although current production is small.

Among commercially mature technologies, advantage for indirect coal liquefaction over direct coal liquefaction are reported by Williams and Larson (2003).

Estimates are reported for sites in China where break-even cost for coal liquefaction may be in the range between 25 to 35 USD/barrel of oil.'

Intensive research and project developments have been implemented from 2001. The World CTL Award is granted to personalities having brought eminent contribution to the understanding and development of Coal liquefaction. The 2009 presentation ceremony will take place in Washington DC (USA) at the World CTL 2009 Conference (25-27 March, 2009).

Coal as a traded commodity

The price of coal has gone up from around $30 per short ton in 2000 to around $123.50 per short ton as of June 25th, 2008.

In North America, a Central Appalachian coal futures contract is currently traded on the New York Mercantile Exchange (trading symbol *QL*). The trading unit is 1,550 short tons per contract, and is quoted in U.S. dollars and cents per ton. Since coal is the principal fuel for generating electricity in the United States, the futures contract provides coal producers and the electric power industry an important tool for hedging and risk management.

In addition to the NYMEX contract, the IntercontinentalExchange (ICE) has European (Rotterdam) and South African (Richards Bay) coal futures available for trading. The trading unit for these contracts is 5,000 tons, and are also quoted in U.S. dollars and cents per ton.

Cultural usage

Coal is the official state mineral of Kentucky and the official state rock of Utah. Both U.S. states have a historic link to coal mining.

Some cultures uphold that children who misbehave will receive coal from Santa Claus for Christmas in their stockings instead of presents.

It is also customary and lucky in Scotland to give coal as a gift on New Year's Day. It happens as part of First-Footing and represents warmth for the year to come.

Environmental effects

There are a number of adverse environmental effects of coal mining and burning.

These effects include:

- release of carbon dioxide and methane, both of which are greenhouse gases, which are causing climate change and global warming according to the IPCC. Coal is the largest contributor to the human-made increase of CO2 in the air.
- waste products including uranium, thorium, and other heavy metals
- acid rain

- interference with groundwater and water table levels
- impact of water use on flows of rivers and consequential impact on other land-uses
- dust nuisance
- subsidence above tunnels, sometimes damaging infrastructure
- rendering land unfit for other uses.
- coal-fired power plants without effective fly ash capture are one of the largest sources of human-caused background radiation exposure.

Energy density

The energy density of coal, i.e. its heating value, is roughly 24 megajoules per kilogram.

The energy density of coal can also be expressed in kilowatt-hours for some unit of mass, the units that electricity is most commonly sold in, to estimate how much coal is required to power electrical appliances. The energy density of coal is 6.67 kW·h/kg and the typical thermodynamic efficiency of coal power plants is about 30%. Of the 6.67 kW·h of energy per kilogram of coal, about 30% of that can successfully be turned into electricity—the rest is waste heat. Coal power plants obtain approximately 2.0 kW·h per kg of burned coal.

As an example, running one 100 watt computer for one year requires 876 kW·h (100 W × 24 h/day × 365 {days in a year} = 876000 W·h = 876 kW·h). Converting this power usage into physical coal consumption:

$$\frac{876\ \mathrm{kW}\cdot\mathrm{h}}{2.0\ \mathrm{kW}\cdot\mathrm{h/kg}} = 438\ \mathrm{kg\ of\ coal} = 966\ \mathrm{pounds\ of\ coal}$$

It takes 438 kg (966 pounds) of coal to power a computer for one full year. One should also take into account transmission and distribution losses caused by resistance and heating in the power lines, which is in the order of 5–10%, depending on distance from the power station and other factors.

Relative carbon cost

Because coal is at least 50% carbon (by mass), then 1 kg of coal contains at least 0.5 kg of carbon, which is

$\frac{0.5\ \mathrm{kg}}{12\cdot\mathrm{kg/kmol}} = \frac{1}{24}\ \mathrm{kmol}$ where 1 mol is equal to N_A (*Avogadro Number*) particles.

This combines with oxygen in the atmosphere during combustion, producing carbon dioxide, with an atomic weight of (12 + 16 × 2 = mass(CO_2) = 44 kg/kmol), so $^1/_{24}$ kmol of CO_2 is produced from the $^1/_{24}$ kmol present in every kilogram of coal, which once trapped in CO_2 weighs approximately

$$\frac{1}{24}\ \mathrm{kmol}\cdot\frac{44\ \mathrm{kg}}{\mathrm{kmol}}=\frac{11}{6}\ \mathrm{kg}\approx 1.83\ \mathrm{kg}\cdot$$

This can be used to put a carbon-cost of energy on the use of coal power. Since the useful energy output of coal is about 30% of the 6.67 kW·h/kg(coal), we can say about 2 kW·h/kg(coal) of energy is produced. Since 1 kg coal roughly translates as 1.83 kg of CO_2, we can say that using electricity from coal produces CO_2 at a rate of about 0.915 kg/(kW·h), or about 0.254 kg/MJ.

This estimate compares favourably with the U.S. Energy Information Agency's 1999 report on CO_2 emissions for energy generation, which quotes a specific emission rate of 950 g CO_2/(kW·h). By comparison, generation from oil in the U.S. was 890 g CO_2/(kW·h), while natural gas was 600 g CO_2/(kW·h). Estimates for specific emission from nuclear power, hydro, and wind energy vary, but are about 100 times lower. See environmental effects of nuclear power for estimates.

Coal fires

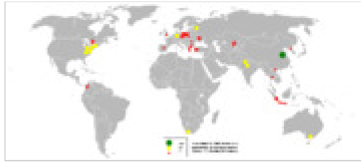
There are hundreds of coal fires burning around the world. Those burning underground can be difficult to locate and many cannot be extinguished. Fires can cause the ground above to subside, their combustion gases are dangerous to life, and breaking out to the surface can initiate surface wildfires. Coal seams can be set on fire by spontaneous combustion or contact with a mine fire or surface fire. A grass fire in a coal area can set dozens of coal seams on fire. Coal fires in China burn 109 million tons of coal a year, emitting 360 million metric tons of CO_2. This contradicts the ratio of 1:1.83 given earlier, but it amounts to 2-3% of the annual worldwide production of CO_2 from fossil fuels, or as much as emitted from all of the cars and light trucks in the United States. In Centralia, Pennsylvania (a borough located in the Coal Region of the United States) an exposed vein of coal ignited in 1962 due to a trash fire in the borough landfill, located in an abandoned anthracite strip mine pit. Attempts to extinguish the fire were unsuccessful, and it continues to burn underground to this day. The Australian Burning Mountain was originally believed to be a volcano, but the smoke and ash comes from a coal fire which may have been burning for over 5,500 years.

At Kuh i Malik in Yagnob Valley, Tajikistan, coal deposits have been burning for thousands of years, creating vast underground labyrinths full of unique minerals, some of them very beautiful. Local people once used this method to mine ammoniac. This place has been well-known since the time of Herodotus, but European geographers mis-interpreted the Ancient Greek descriptions as the evidence of active volcanism in Turkestan (up to the 19th century, when Russian army invaded the area).

The reddish siltstone rock that caps many ridges and buttes in the Powder River Basin (Wyoming), and in western North Dakota is called **porcelanite**, which also may resemble the coal burning waste "clinker" or volcanic " scoria". Clinker is rock that has been fused by the natural burning of coal. In the Powder River Basin approximately 27 to 54 billion tons of coal burned within the past three million years. Wild coal fires in the area were reported by the Lewis and Clark Expedition as well as explorers and settlers in the area.

Production trends

In 2006, China was the top producer of coal with 38% share followed by the USA and India, reports the British Geological Survey.


Coal output in 2005

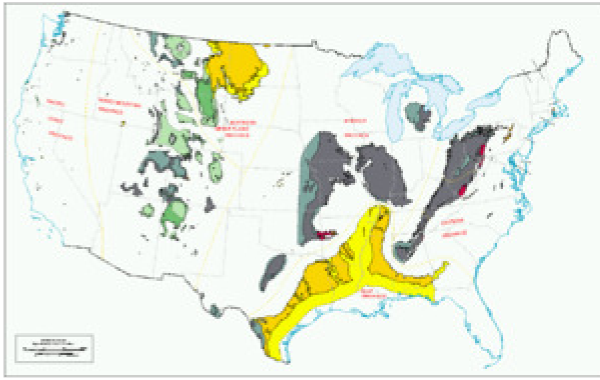
World coal reserves

At the end of 2006 the recoverable coal reserves amounted around 800 or 900 gigatons. The United States Energy Information Administration gives world reserves as 998 billion short tons (equal to 905 gigatons), approximately half of it being hard coal. At the current production rate, this would last 164 years. At the current global total energy consumption of 15 terawatt, there is enough coal to provide the entire planet with all of its energy for 57 years.

The 998 billion tons of recoverable coal reserves estimated by the Energy Information Administration are equal to about 4,417 BBOE (billion barrels of oil equivalent). The amount of coal burned during 2001 was calculated as 2.337 GTOE (gigatonnes of oil equivalent), which is about 46 million barrels of oil equivalent per day. Were consumption to continue at that rate those reserves would last about 263 years. As a comparison, natural gas provided 51 million barrels (oil equivalent), and oil 76 million barrels, per day during 2001.

British Petroleum, in its annual report 2007, estimated at 2006 end, there were 909,064 million tons of *proven* coal reserves worldwide, or 147 years reserves to production ratio. This figure only includes reserves classified as "proven"; exploration drilling programs by mining companies, particularly in under-explored areas, are continually providing new reserves. In many cases, companies are aware of coal deposits that have not been sufficiently drilled to qualify as "proven". However, some nations haven't updated their information and assume reserves remain at the same levels even with withdrawals.

Of the three fossil fuels coal has the most widely distributed reserves; coal is mined in over 100 countries, and on all continents except Antarctica. The largest reserves are found in the USA, Russia, Australia, China, India and South Africa.


US coal regions

Note the table below.

Proved recoverable coal reserves at end-2006 (million tonnes (teragrams))

Country	Bituminous & anthracite	SubBituminous & lignite	TOTAL	Share
USA	111,338	135,305	246,643	27.1
Russia	49,088	107,922	157,010	17.3
China	62,200	52,300	114,500	12.6
India	90,085	2,360	92,445	10.2
Australia	38,600	39,900	78,500	8.6
South Africa	48,750	0	48,750	5.4
Ukraine	16,274	17,879	34,153	3.8
Kazakhstan	28,151	3,128	31,279	3.4
Poland	14,000	0	14,000	1.5
Brazil	0	10,113	10,113	1.1
Germany	183	6,556	6,739	0.7
Colombia	6,230	381	6,611	0.7
Canada	3,471	3,107	6,578	0.7
Czech Republic	2,094	3,458	5,552	0.6
Indonesia	740	4,228	4,968	0.5
Turkey	278	3,908	4,186	0.5
Greece	0	3,900	3,900	0.4
Hungary	198	3,159	3,357	0.4
Pakistan	0	3,050	3,050	0.3
Bulgaria	4	2,183	2,187	0.2
Thailand	0	1,354	1,354	0.1
North Korea	300	300	600	0.1

New Zealand	33	538	571	0.1
Spain	200	330	530	0.1
Zimbabwe	502	0	502	0.1
Romania	22	472	494	0.1
Venezuela	479	0	479	0.1
TOTAL	**478,771**	**430,293**	**909,064**	**100.0**

Major coal producers

Production of Coal by Country and year (million tonnes)

Country	2003	2004	2005	2006
PR China	1722.0	1992.3	2204.7	2380.0
United States	972.3	1008.9	1026.5	1053.6
India	375.4	407.7	428.4	447.3
Australia	351.5	366.1	378.8	373.8
Russian Federation	276.7	281.7	298.5	309.2
South Africa	237.9	243.4	244.4	256.9
Germany	204.9	207.8	202.8	197.2
Indonesia	114.3	132.4	146.9	195.0
Poland	163.8	162.4	159.5	156.1
Total World	**5187.6**	**5585.3**	**5886.7**	**6195.1**

Major coal exporters

Exports of Coal by Country and year (million short tonnes)

Country	2003	2004	2005
Australia	238.1	247.6	257.6
United States	43.0	48.0	49.9
South Africa	78.7	74.9	77.5
CIS (Former Soviet Union)	41.0	55.7	62.3
Poland	16.4	16.3	16.4
Canada	27.7	28.8	31.0
China	103.4	95.5	79.0
South America	57.8	65.9	68.8
Indonesia	107.8	131.4	147.6
Vietnam	N/A	10.3	14.1
Total	**713.9**	**764.0**	**804.2**

Corrosion

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Corrosion means the breaking down of essential properties in a material due to chemical reactions with its surroundings. In the most common use of the word, this means a loss of electrons of metals reacting with water and oxygen. Weakening of iron due to oxidation of the iron atoms is a well-known example of electrochemical corrosion. This is commonly known as rust. This type of damage usually affects metallic materials, and typically produces oxide(s) and/or salt(s) of the original metal. Corrosion also includes the dissolution of ceramic materials and can refer to discoloration and weakening of polymers by the sun's ultraviolet light.

Most structural alloys corrode merely from exposure to moisture in the air, but the process can be strongly affected by exposure to certain substances (see below). Corrosion can be concentrated locally to form a pit or crack, or it can extend across a wide area to produce general deterioration. While some efforts to reduce corrosion merely redirect the damage into less visible, less predictable forms, controlled corrosion treatments such as passivation and chromate-conversion will increase a material's corrosion resistance.

Electrochemical theory


Rust, the most familiar example of corrosion.

One way to understand the structure of metals on the basis of particles is to imagine an array of positively-charged ions sitting in a negatively-charged " gas" of free electrons. Coulombic attraction holds these oppositely-charged particles together, but the positively-charged ions are attracted to negatively charged particles outside the metal as well, such as the negative ions (anions) in an electrolyte. For a given ion at the surface of a metal, there is a certain amount of energy to be gained or lost by dissolving into the electrolyte or becoming a part of the metal, which reflects an atom-scale tug-of-war between the electron gas and dissolved anions. The quantity of energy then strongly depends on a host of variables, including the types of ions in a solution and their concentrations, and the number of electrons present at the metal's surface. In turn, corrosion processes cause electrochemical changes, meaning that they strongly affect all of these variables. The overall interaction between electrons and ions tends to produce a state of local thermodynamic equilibrium that can often be described using basic chemistry and a knowledge of initial conditions.

Galvanic corrosion

Galvanic corrosion occurs when two different metals electrically contact each other and are immersed in an electrolyte. In order for galvanic corrosion to occur, an electrically conductive path and an ionically conductive path are necessary. This effects a galvanic couple where the more active metal corrodes at an accelerated rate and the more noble metal corrodes at a retarded rate. When immersed, neither metal would normally corrode as quickly without the electrically conductive connection (usually via a wire or direct contact). Galvanic corrosion is often utilised in sacrificial anodes. What type of metal(s) to use is readily determined by following the galvanic series. For example, zinc is often used as a sacrificial anode for steel structures, such as pipelines or docked naval ships. Galvanic corrosion is of major interest to the marine industry and also anywhere water can contact pipes or metal structures.

Factors such as relative size of anode (smaller is generally less desirable), types of metal, and operating conditions (temperature, humidity, salinity, *&c.*) will affect galvanic corrosion. The surface area ratio of the anode and cathode will directly affect the corrosion rates of the materials.

Galvanic series

In a given saeeff environment (one standard medium is aerated, room-temperature seawater), one metal will be either more *noble* or more *active* than the next, based on how strongly its ions are bound to the surface. Two metals in electrical contact share the same electron gas, so that the tug-of-war at each surface is translated into a competition for free electrons between the two materials. The noble metal will tend to take electrons from the active one, while the electrolyte hosts a flow of ions in the same direction. The resulting mass flow or electrical current can be measured to establish a hierarchy of materials in the medium of interest. This hierarchy is called a *Galvanic series*, and can be a very useful in predicting and understanding corrosion.

Resistance to corrosion

Some metals are more intrinsically resistant to corrosion than others, either due to the fundamental nature of the electrochemical processes involved or due to the details of how reaction products form. For some examples, see galvanic series.. If a more susceptible material is used, many techniques can be applied during an item's manufacture and use to protect its materials from damage.

Intrinsic chemistry


Gold nuggets do not naturally corrode, even on a geological time scale.

The materials most resistant to corrosion are those for which corrosion is thermodynamically unfavorable. Any corrosion products of gold or platinum tend to decompose spontaneously into pure metal, which is why these elements can be found in metallic form on Earth, and is a large part of their intrinsic value. More common "base" metals can only be protected by more temporary means.

Some metals have naturally slow reaction kinetics, even though their corrosion is thermodynamically favorable. These include such metals as zinc, magnesium, and cadmium. While corrosion of these metals is continuous and ongoing, it happens at an acceptably slow rate. An extreme example is graphite, which releases large amounts of energy upon oxidation, but has such slow kinetics that it is effectively immune to electrochemical corrosion under normal conditions.

Passivation

Given the right conditions, a thin film of corrosion products can form on a metal's surface spontaneously, acting as a barrier to further oxidation. When this layer stops growing at less than a micrometre thick under the conditions that a material will be used in, the phenomenon is known as passivation (rust, for example, usually grows to be much thicker, and so is not considered passivation, because this mixed oxidized layer is not protective). While this effect is in some sense a property of the material, it serves as an indirect kinetic barrier: the reaction is often quite rapid unless and until an impermeable layer forms. Passivation in air and water at moderate pH is seen in such materials as aluminium, stainless steel, titanium, and silicon.

These conditions required for passivation are specific to the material. The effect of pH is recorded using Pourbaix diagrams, but many other factors are influential. Some conditions that inhibit passivation include: high pH for aluminium, low pH or the presence of chloride ions for stainless steel, high temperature for titanium (in which case the oxide dissolves into the metal, rather than the electrolyte) and fluoride ions for silicon. On the other hand, sometimes unusual conditions can bring on passivation in materials that are normally unprotected, as the alkaline environment of concrete does for steel rebar. Exposure to a liquid metal such as mercury or hot solder can often circumvent passivation mechanisms.

Corrosion in passivated materials

Passivation is extremely useful in alleviating corrosion damage, but care must be taken not to trust it too thoroughly. Even a high-quality alloy will corrode if its ability to form a passivating film is hindered. Because the resulting modes of corrosion are more exotic and their immediate results are less visible than rust and other bulk corrosion, they often escape notice and cause problems among those who are not familiar with them.

Pitting corrosion

Certain conditions, such as low concentrations of oxygen or high concentrations of species such as chloride which compete as anions, can interfere with a given alloy's ability to re-form a passivating film. In the worst case, almost all of the surface will remain protected, but tiny local fluctuations will degrade the oxide film in a few critical points. Corrosion at these points will be greatly amplified, and can cause *corrosion pits* of several types, depending upon conditions. While the corrosion pits only nucleate under fairly extreme circumstances, they can continue to grow even when conditions return to normal, since the interior of a pit is naturally deprived of oxygen. In extreme cases, the sharp tips of extremely long and narrow can cause stress concentration to the point that otherwise tough alloys can shatter, or a thin film pierced by an invisibly small hole can hide a thumb sized pit from view. These problems are especially dangerous because they are difficult to detect before a part or structure fails. Pitting remains among the most common and damaging forms of corrosion in passivated alloys, but it can be prevented by control of the alloy's environment, which often includes ensuring that the material is exposed to oxygen uniformly (i.e., eliminating crevices).and many rocks can be reformed from corrosion.

Weld decay and knifeline attack

Stainless steel can pose special corrosion challenges, since its passivating behaviour relies on the presence of a minor alloying component (Chromium, typically only 18%). Due to the elevated temperatures of welding or during improper heat treatment, chromium carbides can form in the grain boundaries of stainless alloys. This chemical reaction robs the material of chromium in the zone near the grain boundary, making those areas much less resistant to corrosion. This creates a galvanic couple with the well-protected alloy nearby, which leads to *weld decay* (corrosion of the grain boundaries near welds) in highly corrosive environments. Special alloys, either with low carbon content or with added carbon " getters" such as titanium and niobium (in types 321 and 347, respectively), can prevent this effect, but the latter require special heat treatment after welding to prevent the similar phenomenon of *knifeline attack*. As its name applies, this is limited to a small zone, often only a few micrometres across, which causes it to proceed more rapidly. This zone is very near the weld, making it even less noticeable[1].

Crevice corrosion

Crevice corrosion is a corrosion occurring in spaces to which the access of the working fluid from the environment is limited. These spaces are generally called crevices. Examples of crevices are gaps and contact areas between parts, under gaskets or seals, inside cracks and seams, spaces filled with deposits and under sludge piles.

Microbial corrosion

Microbial corrosion, or bacterial corrosion, is a corrosion caused or promoted by microorganisms, usually chemoautotrophs. It can apply to both metals and non-metallic materials, in both the presence and lack of oxygen. Sulfate-reducing bacteria are common in lack of oxygen; they produce hydrogen sulfide, causing sulfide stress cracking. In presence of oxygen, some bacteria directly oxidize iron to iron oxides and hydroxides, other bacteria oxidize sulfur and produce sulfuric acid causing biogenic sulfide corrosion. Concentration cells can form in the deposits of corrosion products, causing and enhancing galvanic corrosion.

Titanic's bow exhibiting microbial corrosion damage in the form of 'rusticles'

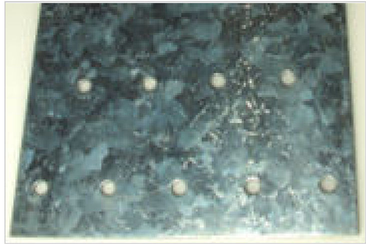
High temperature corrosion

High temperature corrosion is chemical deterioration of a material (typically a metal) under very high temperature conditions. This non-galvanic form of corrosion can occur when a metal is subject to a high temperature atmosphere containing oxygen, sulfur or other compounds capable of oxidising (or assisting the oxidation of) the material concerned. For example, materials used in aerospace, power generation and even in car engines have to resist sustained periods at high temperature in which they may be exposed to an atmosphere containing potentially highly corrosive products of combustion.

The products of high temperature corrosion can potentially be turned to the advantage of the engineer. The formation of oxides on stainless steels, for example, can provide a protective layer preventing further atmospheric attack, allowing for a material to be used for sustained periods at both room and high temperature in hostile conditions. Such high temperature corrosion products in the form of compacted oxide layer glazes have also been shown to prevent or reduce wear during high temperature sliding contact of metallic (or metallic and ceramic) surfaces.

Surface treatments

Applied coatings



Galvanized surface

Plating, painting, and the application of enamel are the most common anti-corrosion treatments. They work by providing a barrier of corrosion-resistant material between the damaging environment and the (often cheaper, tougher, and/or easier-to-process) structural material. Aside from cosmetic and manufacturing issues, there are tradeoffs in mechanical flexibility versus resistance to abrasion and high temperature. Platings usually fail only in small sections, and if the plating is more noble than the substrate (for example, chromium on steel), a galvanic couple will cause any exposed area to corrode much more rapidly than an unplated surface would. For this reason, it is often wise to plate with a more active metal such as zinc or cadmium.

Reactive coatings

If the environment is controlled (especially in recirculating systems), corrosion inhibitors can often be added to it. These form an electrically insulating and/or chemically impermeable coating on exposed metal surfaces, to suppress electrochemical reactions. Such methods obviously make the system less sensitive to scratches or defects in the coating, since extra inhibitors can be made available wherever metal becomes exposed. Chemicals that inhibit corrosion include some of the salts in hard water (Roman water systems are famous for their mineral deposits), chromates, phosphates, and a wide range of specially-designed chemicals that resemble surfactants (i.e. long-chain organic molecules with ionic end groups).

This figure-8 descender is annodized with a yellow finish. Climbing equipment is available in a wide range of anodized colors.

Anodization

Aluminium alloys often undergo a surface treatment. Electrochemical conditions in the bath are carefully adjusted so that uniform pores several nanometers wide appear in the metal's oxide film. These pores allow the oxide to grow much thicker than passivating conditions would allow. At the end of the treatment, the pores are allowed to seal, forming a harder-than-usual surface layer. If this coating is scratched, normal passivation processes take over to protect the damaged area.

Controlled Permeability Formwork

Controlled Permeability Formwork (CPF) is a method of preventing the corrosion of reinforcement by naturally enhancing the durability of the cover during concrete placement. CPF has been used in environments to combat the effects of Carbonation, chlorides, frost and abrasion.

Cathodic protection

Cathodic protection (CP) is a technique to control the corrosion of a metal surface by making that surface the cathode of an electrochemical cell.

It is a method used to protect metal structures from corrosion. Cathodic protection systems are most commonly used to protect steel, water, and fuel pipelines and tanks; steel pier piles, ships, and offshore oil platforms.

For effective CP, the potential of the steel surface is polarized (pushed) more negative until the metal surface has a uniform potential. With a uniform potential, the driving force for the corrosion reaction is halted. For galvanic CP systems, the anode material corrodes under the influence of the steel, and eventually it must be replaced. The polarization is caused by the current flow from the anode to the cathode, driven by the difference in electrochemical potential between the anode and the cathode.

For larger structures, galvanic anodes cannot economically deliver enough current to provide complete protection. Impressed Current Cathodic Protection (ICCP) systems use anodes connected to a DC power source (a cathodic protection rectifier). Anodes for ICCP systems are tubular and solid rod shapes of various specialized materials. These include high silicon cast iron, graphite, mixed metal oxide or platinum coated titanium or niobium coated rod and wires.

Economic impact

The US Federal Highway Administration released a study, entitled *Corrosion Costs and Preventive Strategies in the United States,* in 2002 on the direct costs associated with metallic corrosion in nearly every U.S. industry sector. The study showed that for 1998 the total annual estimated direct cost of corrosion in the U.S. was approximately $276 billion (approximately 3.1% of the US gross domestic product). FHWA Report Number:FHWA-RD-01-156. The NACE International website has a summary slideshow of the report findings. Jones[1] writes that electrochemical corrosion causes between $8 billion and $128 billion in economic damage per year in the United States alone, degrading structures, machines, and containers.


The collapsed Silver Bridge, as seen from the Ohio side

Rust is one of the most common causes of bridge accidents for example. As rust has a much higher volume than the originating mass of iron, its build-up can also cause failure by forcing apart adjacent parts. It was the cause of the collapse of the Mianus river bridge in 1983, when the bearings rusted internally and pushed one corner of the road slab off its support. Three drivers on the roadway at the time died as the slab fell into the river below. The following NTSB investigation showed that a drain in the road had been blocked for road re-surfacing, and had not been unblocked so that runoff water penetrated the support hangers. It was also difficult for maintenance engineers to see the bearings from the inspection walkway. Rust was also an important factor in the Silver Bridge disaster of 1967 in West Virginia, when a steel suspension bridge collapsed in less than a minute, killing 46 drivers and passengers on the bridge at the time.

Similarly corrosion of concrete-covered steel and iron can cause the concrete to spall, creating severe structural problems. It is one of the most common failure modes of reinforced concrete bridges.

Corrosion in nonmetals

Most ceramic materials are almost entirely immune to corrosion. The strong ionic and/or covalent bonds that hold them together leave very little free chemical energy in the structure; they can be thought of as already corroded. When corrosion does occur, it is almost always a simple dissolution of the material or chemical reaction, rather than an electrochemical process. A common example of corrosion protection in ceramics is the lime added to soda-lime glass to reduce its solubility in water; though it is not nearly as soluble as pure sodium silicate, normal glass does form sub-microscopic flaws when exposed to moisture. Due to its brittleness, such flaws cause a dramatic reduction in the strength of a glass object during its first few hours at room temperature.

Polymer degradation is due to a wide array of complex and often poorly-understood physiochemical processes. These are strikingly different from the other processes discussed here, and so the term "corrosion" is only applied to them in a loose sense of the word. Because of their large molecular weight, very little entropy can be gained by mixing a given mass of polymer with another substance, making them generally quite difficult to dissolve. While dissolution is a problem in some polymer applications, it is relatively simple to design against. A more common and related problem is *swelling*, where small molecules infiltrate the structure, reducing strength and stiffness and causing a volume change. Conversely, many polymers (notably flexible vinyl) are intentionally swelled with plasticizers, which can be leached out of the structure, causing brittleness or other undesirable changes. The most common form of degradation, however, is a decrease in polymer chain length. Mechanisms which break polymer chains are familiar to biologists because of their effect on DNA: ionizing radiation (most commonly ultraviolet light), free radicals, and oxidizers such as oxygen, ozone, and chlorine. Additives can slow these process very effectively, and can be as simple as a UV-absorbing pigment (i.e., titanium dioxide or carbon black). Plastic shopping bags often do not include these additives so that they break down more easily as litter.

Corrosion of glasses

The corrosion of silicate glasses in aqueous solutions is governed by two mechanisms: diffusion-controlled leaching (ion exchange) and glass network hydrolytic dissolution. Both corrosion mechanisms strongly depend on the pH of contacting solution: the rate of ion exchange decreases with pH as $10^{-0.5pH}$ whereas the rate of hydrolytic dissolution increases with pH as $10^{0.5pH}$

Numerically, corrosion rates of glasses are characterised by normalised corrosion rates of elements NR_i (g/cm^2 d) which are determined as the ratio of total amount of released species into the water M_i (g) to the water-contacting surface area S (cm^2), time of contact t (days) and weight fraction content of the element in the glass f_i:

$NR_i = M_i / S \cdot f_i \cdot t$

The overall corrosion rate is a sum of contributions from both mechanisms (leaching + dissolution) $NR_i = Nrx_i + NRh$. Diffusion-controlled leaching (ion exchange) is characteristic of the initial phase of corrosion and involves replacement of alkali ions in the glass by a hydronium (H_3O^+) ion from the solution. It causes an ion-selective depletion of near surface layers of glasses and gives an inverse square root dependence of corrosion rate with exposure time. The diffusion controlled normalised leaching rate of cations from glasses (g/cm^2 d) is given by:

$NRx_i=2\cdot\rho\cdot(D_i/\pi\cdot t)^{1/2}$

where t is time, D_i is the i-th cation effective diffusion coefficient (cm^2/d), which depends on pH of contacting water as $D_i = D_{i0}\cdot 10^{-pH}$, and ρ is the density of the glass (g/cm^3).

Glass network dissolution is characteristic of the later phases of corrosion and causes a congruent release of ions into the water solution at a time-independent rate in dilute solutions (g/cm^2 d):

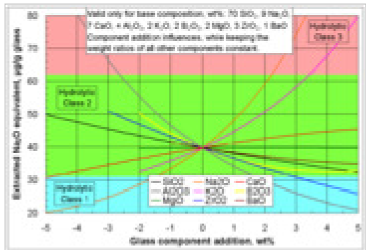
$NRh=\rho r_h$,

where r_h is the stationary hydrolysis (dissolution) rate of the glass (cm/d). In closed systems the consumption of protons from the aqueous phase increases the pH and causes a fast transition to hydrolysis. However further silica saturation of solution impedes hydrolysis and causes the glass to return to an ion-exchange, e.g. diffusion-controlled regime of corrosion.

In typical natural conditions normalised corrosion rates of silicate glasses are very low and are of the order of 10^{-7} - 10^{-5} g/cm^2 d. The very high durability of silicate glasses in water makes them suitable for hazardous and nuclear waste immobilisation.

Glass corrosion tests

There exist numerous standardized procedures for measuring the corrosion (also called **chemical durability**) of glasses in neutral, basic, and acidic environments, under simulated environmental conditions, in simulated body fluid, at high temperature and pressure, and under other conditions.



Influences of selected glass component additions on the chemical durability against water corrosion of a specific base glass (corrosion test ISO 719).

In the standard procedure ISO 719 a test of the extraction of water soluble basic compounds under neutral conditions is described: 2 g glass, particle size 300-500 µm, is kept for 60 min in 50 ml de-ionized water of grade 2 at 98°C. 25 ml of the obtained solution is titrated against 0.01 mol/l HCl solution. The volume of HCl needed for neutralization is recorded and classified following the values in the table below.

0.01M HCl needed to neutralize extracted basic oxides, ml	Extracted Na_2O equivalent, µg	Hydrolytic class
to 0.1	to 31	1
above 0.1 to 0.2	above 31 to 62	2

above 0.2 to 0.85	above 62 to 264	3
above 0.85 to 2.0	above 264 to 620	4
above 2.0 to 3.5	above 620 to 1085	5
above 3.5	above 1085	>5

Cutty Sark

2008/9 Schools Wikipedia Selection. Related subjects: Air & Sea transport

Template:Infobox Ship

The ***Cutty Sark*** is a clipper ship. Built in 1869, she served as a merchant vessel (the last clipper to be built for that purpose), and then as a training ship until being put on public display in 1954. She is preserved in dry dock at Greenwich in London, but was damaged in a fire on 21 May 2007 while undergoing extensive restoration.

Etymology

The ship is named after the *cutty sark* (Scots: a short chemise or undergarment). This was the nickname of the fictional character Nannie (also the name of the ship's figurehead) in Robert Burns' 1791 comic poem *Tam o' Shanter*. She was wearing a linen *cutty sark* that she had been given as a child, therefore it was far too small for her. The erotic sight of her dancing in such a short undergarment caused Tam to cry out "Weel done, Cutty-sark", which subsequently became a well known idiom.

History


Cutty Sark sailing

She was designed by Hercules Linton and built in 1869 at Dumbarton, Scotland, by the firm of Scott & Linton, for Captain John "Jock" "White Hat" Willis; Scott & Linton was liquidated, and she was launched November 22 of that year by William Denny & Brothers.

Cutty Sark was destined for the tea trade, then an intensely competitive race across the globe from China to London, with immense profits to the ship to arrive with the first tea of the year. However, she did not distinguish herself; in the most famous race, against *Thermopylae* in 1872, both ships left Shanghai together on June 18, but two weeks later *Cutty Sark* lost her rudder after passing through the Sunda Strait, and arrived in London on October 18, a week after *Thermopylae*, a total passage of 122 days. Her legendary reputation is supported by the fact that her captain chose to continue this race with an improvised rudder instead of putting into port for a replacement, yet was only beaten by one week.

In the end, clippers lost out to steamships, which could pass through the recently-opened Suez Canal and deliver goods more reliably, if not quite so quickly, which proved to be better for business. *Cutty Sark* was then used on the Australian wool trade. Under the respected Captain Richard Woodget, she did very well, posting Australia-to-Britain times of as little as 67 days. Her best run, 360 nautical miles (666 km) in 24 hours (an average 15kt, 27.75 km/h), was said to have been the fastest of any ship of her size.

In 1895 Willis sold her to the Portuguese firm Ferreira and she was renamed ***Ferreira*** after the firm, although her crews referred to her as ***Pequena Camisola*** ("little shirt", a straight translation of the Scots "cutty sark"). In 1916 she was dismasted off the Cape of Good Hope, sold, re-rigged in Cape Town as a barquentine, and renamed ***Maria do Amparo***. In 1922 she was bought by Captain Wilfred Dowman, who restored her to her original appearance and used her as a stationary training ship. In 1954 she was moved to a custom-built dry-dock at Greenwich.

Cutty Sark is also preserved in literature in Hart Crane's long poem "The Bridge" which was published in 1930.

Museum ship

The *Cutty Sark* was preserved as a museum ship and popular tourist attraction. She is located near the centre of Greenwich, in south-east London, close aboard the National Maritime Museum, the former Greenwich Hospital, and Greenwich Park. She is also a prominent landmark on the route of the London Marathon. She usually flies signal flags from her ensign halyard reading "JKWS", which is the code representing Cutty Sark in the International Code of Signals, introduced in 1857.

Cutty Sark in Greenwich, October 2003

The ship is in the care of the Cutty Sark Trust, whose president, the Duke of Edinburgh, was instrumental in ensuring her preservation, when he set up the Cutty Sark Society in 1951. The Trust replaced the Society in 2000. She is a Grade I listed monument and is on the Buildings At Risk Register.

Cutty Sark station on the Docklands Light Railway is one minute's walk away, with connections to central London and the London Underground. Greenwich Pier is next to the ship, and is served by scheduled river boats from piers in central London. A tourist information office stands to the east of the ship.

Conservation and fire

On the morning of 21 May 2007 the *Cutty Sark*, which had been closed and partly dismantled for conservation work, caught fire, and burned for several hours before the London Fire Brigade could bring the fire under control. Initial reports indicated that that damage was extensive, with most of the wooden structure in the centre having been lost.

In an interview the next day, Richard Doughty, the chief executive of the Cutty Sark Trust revealed that at least half of the "fabric" (timbers, etc) of the ship had not been on site as it had been removed during the preservation work. Doughty expressed that the trust was most worried about the state of iron framework to which the fabric was attached. He did not know how much more the ship would cost to restore, but estimated it at an additional £5–10 million, bringing the total cost of the ship's restoration to £30–35 million.

Aerial video footage showed extensive damage, but seemed to indicate that the ship had not been destroyed in its entirety. A fire officer present at the scene said in a BBC interview that when they arrived, there had been "a well-developed fire throughout the ship". The bow section looked to be relatively unscathed and the stern also appeared to have survived without major damage. The fire seemed to have been concentrated in the centre of the ship. The chairman of Cutty Sark Enterprises said after inspecting the site: "The decks are unsalvageable but around 50% of the planking had already been removed; however, the damage is not as bad as originally expected."

As part of the restoration work planned before the fire, it was proposed that the ship be raised three metres, to allow the construction of a state of the art museum space beneath. This would allow visitors to view her from below.

For a long time, there had been been growing criticism of the policies of the Cutty Sark Trust and its stance that the most important thing was to preserve as much as possible of the original fabric. The fire damage has been put forth as a reason for the Cutty Sark to be rebuilt in a manner that would allow her to put to sea again by proponents of the idea. However, the Cutty Sark Trust have found that less than 5% of the original fabric was lost in the fire, as the decks which were destroyed were non-original additions. There are currently two petitions to the UK Prime Minister, one for funds to restore the ship, and the other for funds to restore the ship into commission as a sail training vessel.

In addition to explaining how and why the ship is being saved, the exhibition features a new film presentation, a re-creation of the master's saloon, and interactive exhibits about the project. Live webcam views of the conservation work allow the visitor to see remotely the work being carried out on the ship.

General specifications

The *Cutty Sark* is one of only three surviving ships of its time that has a composite wrought iron frame structure covered by wooden planking. The hull has a Muntz metal coating.

- Tonnage: 921 tons (2,608 m³)
- Hull length: 212.5 ft (64.8 m)
- Beam: 36 ft (11 m)
- Draft: 21 ft (6.4 m)


Bow of the *Cutty Sark*

Yard lengths (after being cut down in Sydney harbour):

- Fore
  - fore course 21.0 yd (19.2 m)
  - lower topsail 16.8 yd (15.4 m)
  - upper topsail 14.6 yd (13.4 m)
  - topgallant 11.5 yd (10.5 m)
  - royal 9.4 yd (8.6 m)
- Main
  - main course 21.6 yd (19.8 m)
  - lower topsail 18.5 yd (16.9 m)
  - upper topsail 16.8 yd (15.4 m)
  - topgallant 14.2 yd (13.0 m)
  - royal 10.4 yd (9.5 m)
- Mizzen
  - mizzen course 17.4 yd (15.9 m)
  - lower topsail 14.9 yd (13.6 m)
  - upper topsail 13.4 yd (12.3 m)
  - topgallant 11.0 yd (10.1 m)
  - royal 8.2 yd (7.5 m)
  - spanker 14.1 yd (12.9 m)

Dam

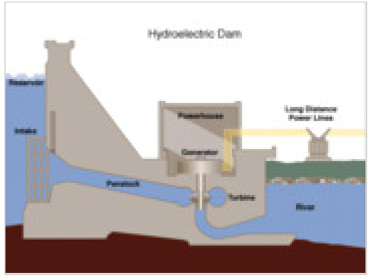
2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Karun-3 dam, Iran.

A **dam** is a barrier that divides waters. Dams generally serve the primary purpose of retaining water, while other structures such as floodgates, levees, and dikes are used to prevent water flow into specific land regions. The tallest dam in the world is the 300 meter high Nurek Dam in Tajikistan.

History

The word *dam* can be traced back to Middle English, and before that, from Middle Dutch, as seen in the names of many old cities.



Hydroelectric dam in cross section.

Some of the grandest and largest dams were constructed in Ceylon (Sri Lanka). Dams in Yodha Wewa and Parakrama Samudra of Sri Lanka were the largest until the 20th Century . As per Needham, Abhaya Wewa is the oldest reservoir that was made by the use of a dam, which has been dated to 300 BC. (Most of the first Dams were built in Mesopotamia up to 7,000 years ago. These were used to control the water level, for Mesopotamia's weather affected the Tigris and Euphrates rivers, and could be quite unpredictable. The earliest recorded dam is believed to have been on the Sadd Al-Kafara at Wadi Al-Garawi, which is located about 25 kilometers south of Cairo, and built around 2600 B.C. It was destroyed by heavy rain shortly afterwards. The Romans were also great dam builders, with many examples such as the three dams at Subiaco on the river Anio in Italy. Many large dams also survive at Merida in Spain.

The oldest surviving and standing dam in the world is believed to be the Grand Anicut, also known as the Kallanai, an ancient dam built on the Kaveri River in the state of Tamil Nadu located in southern India. It was built by the Chola king Karikalan, and dates back to the 2nd Century AD.

The Kallanai is a massive dam of unhewn stone, over 300 meters long, 4.5 meters high and 20 meters (60 ft) wide, across the main stream of the Kaveri. The purpose of the dam was to divert the waters of the Cauvery across the fertile Delta region for irrigation via canals. The dam is still in excellent repair, and served as a model for later engineers, including the Sir Arthur Cotton's 19th-century dam across the Kollidam, the major tributary of the Cauvery. The land area irrigated by the ancient irrigation network, of which the dam was the centerpiece, was 69,000 acres (280 square kilometers). By the early 20th Century the irrigated area had been increased to about 1,000,000 acres (4,000 square kilometers).

Du Jiang Yan in China is the oldest surviving irrigation system included a dam that directed waterflow. It was finished in 251 B.C.. In ancient China, the Prime Minister of Chu (state), Sunshu Ao, is the first known hydraulic engineer of China. He served Duke Zhuang of Chu during the reign of King Ding of Zhou (606

BC-586 BC), ruler of the Eastern Zhou Dynasty. His large earthen dam flooded a valley in modern-day northern Anhui province that created an enormous irrigation reservoir (62 miles in circumference), a reservoir that is still present today.

In the Netherlands, a low-lying country, *dams* were often applied to block rivers in order to regulate the water level and to prevent the sea from entering the marsh lands. Such dams often marked the beginning of a town or city because it was easy to cross the river at such a place, and often gave rise to the respective place's names in Dutch. For instance the Dutch capital Amsterdam (old name Amstelredam) started with a *dam* through the river Amstel in the late 12th Century , and Rotterdam started with a *dam* through the river Rotte, a minor tributary of the Nieuwe Maas. The central square of Amsterdam, believed to be the original place of the 800 year old dam, still carries the name *Dam Square* or simply *the Dam*.

Types of dams

Dams can be formed by human agency, natural causes, or even by the intervention of wildlife such as beavers. Man-made dams are typically classified according to their size (height), intended purpose or structure.

By size

International standards define *large dams* as higher than 15 meters and *major dams* as over 150 meters in height.

By purpose

Intended purposes include providing water for irrigation or town or city water supply, improving navigation, creating a reservoir of water to supply industrial uses, generating hydroelectric power, creating recreation areas or habitat for fish and wildlife, flood control and containing effluent from industrial sites such as mines or factories. Few dams serve all of these purposes but some multi-purpose dams serve more than one.

A *saddle dam* is an auxiliary dam constructed to confine the reservoir created by a primary dam either to permit a higher water elevation and storage or to limit the extent of a reservoir for increased efficiency. An auxiliary dam is constructed in a low spot or *saddle* through which the reservoir would otherwise escape. On occasion, a reservoir is contained by a similar structure called a dike to prevent inundation of nearby land. Dikes are commonly used for *reclamation* of arable land from a shallow lake. This is similar to a levee, which is a wall or embankment built along a river or stream to protect adjacent land from flooding.

An *overflow dam* is designed to be over topped. A weir is a type of small overflow dam that can be used for flow measurement.

A *check dam* is a small dam designed to reduce flow velocity and control soil erosion. Conversely, a *wing dam* is a structure that only partly restricts a waterway, creating a faster channel that resists the accumulation of sediment.

A *dry dam* is a dam designed to control flooding. It normally holds back no water and allows the channel to flow freely, except during periods of intense flow that would otherwise cause flooding downstream.

A *diversionary dam* is a structure designed to divert all or a portion of the flow of a river from its natural course.

By structure

Based on structure and material used, dams are classified as timber dams, arch-gravity dams, embankment dams or masonry dams, with several subtypes.

Masonry dams

Arch dams

Hoover Dam, a concrete arch-gravity dam in the Black Canyon of the Colorado River.

In the arch dam, stability is obtained by a combination of arch and gravity action. If the upstream face is vertical the entire weight of the dam must be carried to the foundation by gravity, while the distribution of the normal hydrostatic pressure between vertical cantilever and arch action will depend upon the stiffness of the dam in a vertical and horizontal direction. When the upstream face is sloped the distribution is more complicated. The normal component of the weight of the arch ring may be taken by the arch action, while the normal hydrostatic pressure will be distributed as described above. For this type of dam, firm reliable supports at the abutments (either buttress or canyon side wall) are more important. The most desirable place for an arch dam is a narrow canyon with steep side walls composed of sound rock. The safety of an arch dam is dependent on the strength of the side wall abutments, hence not only should the arch be well seated on the side walls but also the character of the rock should be carefully inspected.

Two types of single-arch dams are in use, namely the constant-angle and the constant-radius dam. The constant-radius type employs the same face radius at all elevations of the dam, which means that as the channel grows narrower towards the bottom of the dam the central angle subtended by the face of the dam becomes smaller. Jones Falls Dam, in Canada, is a constant radius dam. In a constant-angle dam, also known as a variable radius dam, this subtended angle is kept a constant and the variation in distance between the abutments at various levels are taken care of by varying the radii. Constant-radius dams are much less common than constant-angle dams. Parker Dam is a constant-angle arch dam.

A similar type is the double-curvature or thin-shell dam. Wildhorse Dam near Mountain City, Nevada in the United States is an example of the type. This method of construction minimizes the amount of concrete necessary for construction but transmits large loads to the foundation and abutments. The appearance is similar to a single-arch dam but with a distinct vertical curvature to it as well lending it the vague appearance of a concave lens as viewed from downstream.

The multiple-arch dam consists of a number of single-arch dams with concrete buttresses as the supporting abutments. The multiple-arch dam does not require as many buttresses as the hollow gravity type, but requires good rock foundation because the buttress loads are heavy.

Gravity dams

In a gravity dam, stability is secured by making it of such a size and shape that it will resist overturning, sliding and crushing at the toe. The dam will not overturn provided that the moment around the turning point, caused by the water pressure is smaller than the moment caused by the weight of the dam. This is the case if the resultant force of water pressure and weight falls within the base of the dam. However, in order to prevent tensile stress at the upstream face and excessive compressive stress at the downstream face, the dam cross section is usually designed so that the resultant falls within the middle at all elevations of the cross section (the core). For this type of dam, impervious foundations with high *bearing* strength are essential.

The Gilboa Dam in the Catskill Mountains of New York State is an example of a "solid" gravity dam.

When situated on a suitable site, a gravity dam inspires more confidence in the layman than any other type; it has mass that lends an atmosphere of permanence, stability, and safety. When built on a carefully studied foundation with stresses calculated from completely evaluated loads, the gravity dam probably represents the best developed example of the art of dam building. This is significant because the fear of flood is a strong motivator in many regions, and has resulted in gravity dams being built in some instances where an arch dam would have been more economical.

Gravity dams are classified as "solid" or "hollow." The solid form is the more widely used of the two, though the hollow dam is frequently more economical to construct. Gravity dams can also be classified as "overflow" (spillway) and "non-overflow." Grand Coulee Dam is a solid gravity dam and Itaipu Dam is a hollow gravity dam. A gravity dam can be combined with an arch dam, an arch-gravity dam, for areas with massive amounts of water flow but less material available for a purely gravity dam.

Arch-gravity dams

Embankment dams

Embankment dams are made from compacted earth, and have two main types, rock-fill and earth-fill dams. Embankment dams rely on their weight to hold back the force of water, like the gravity dams made from concrete.

The San Luis Dam near Los Banos, California is an embankment dam.

Rock-fill dams

Rock-fill dams are embankments of compacted free-draining granular earth with an impervious zone. The earth utilized often contains a large percentage of large particles hence the term *rock-fill*. The impervious zone may be on the upstream face and made of masonry, concrete, plastic membrane, steel sheet piles, timber or other material. The impervious zone may also be within the embankment in which case it is referred to as a *core*. In the instances where clay is utilized as the impervious material the dam is referred to as a *composite* dam. To prevent internal erosion of clay into the rock fill due to seepage forces, the core is separated using a filter. Filters are specifically graded soil designed to prevent the migration of fine grain soil particles. When suitable material is at hand, transportation is minimized leading to cost savings during construction. Rock-fill dams are resistant to damage from earthquakes. However, inadequate quality control during construction can lead to poor compaction and sand in the embankment which can lead to liquefaction of the rock-fill during an earthquake. Liquefaction potential can be reduced by keeping susceptible material from being saturated, and by providing adequate compaction during construction. An example of a rock-fill dam is New Melones Dam in California.

Earth-fill dams

Earth-fill dams, also called earthen, rolled-earth or simply earth dams, are constructed as a simple embankment of well compacted earth. A *homogeneous* rolled-earth dam is entirely constructed of one type of material but may contain a drain layer to collect *seep* water. A *zoned-earth* dam has distinct parts or *zones* of dissimilar material, typically a locally plentiful *shell* with a watertight clay core. Modern zoned-earth embankments employ filter and drain zones to collect and remove seep water and preserve the integrity of the downstream shell zone. An outdated method of zoned earth dam construction utilized a hydraulic fill to produce a watertight core. *Rolled-earth* dams may also employ a watertight facing or core in the manner of a rock-fill dam. An interesting type of temporary earth dam occasionally used in high latitudes is the *frozen-core* dam, in which a coolant is circulated through pipes inside the dam to maintain a watertight region of permafrost within it.

Because earthen dams can be constructed from materials found on-site or nearby, they can be very cost-effective in regions where the cost of producing or bringing in concrete would be prohibitive. This makes it better for the environment too.

Asphalt-Concrete Core

A third type of embankment dam is built with asphalt concrete core. The majority of such dams are built with rock and/or gravel as the main fill material. Almost 100 dams of this design have now been built world-wide since the first such dam was completed in 1962. All asphalt-concrete core dams built so far have an excellent performance record. The type of asphalt used is a viscoelastic-plastic material that can adjust to the movements and deformations imposed on the embankment as a whole, and to settlements in the foundation. The flexible properties of the asphalt make such dams especially suited in earthquake regions.

Cofferdams

A cofferdam is a (usually temporary) barrier constructed to exclude water from an area that is normally submerged. Made commonly of wood, concrete or steel sheet piling, cofferdams are used to allow construction on the foundation of permanent dams, bridges, and similar structures. When the project is completed, the cofferdam may be demolished or removed. See also causeway and retaining wall. Common uses for cofferdams include construction and repair of off shore oil platforms. In such cases the cofferdam is fabricated from sheet steel and welded into place under water. Air is pumped into the space, displacing the water allowing a dry work environment below the surface. Upon completion the cofferdam is usually deconstructed unless the area requires continuous maintenance.

A cofferdam during the construction of locks at the Montgomery Point Lock and Dam.

Timber dams

Timber dams were widely used in the early part of the industrial revolution and in frontier areas due to ease and speed of construction. Rarely built in modern times by humans due to relatively short lifespan and limited height to which they can be built, timber dams must be kept constantly wet in order to maintain their water retention properties and limit deterioration by rot, similar to a barrel. The locations where timber dams are most economical to build are those where timber is plentiful, cement is costly or difficult to transport, and either a low head diversion dam is required or longevity is not an issue. Timber dams were once numerous, especially in the North American west, but most have failed, been hidden under earth embankments or been replaced with entirely new structures. Two common variations of timber dams were the *crib* and the *plank*.

A timber crib dam in Michigan, photographed in 1978.

Timber crib dams were erected of heavy timbers or dressed logs in the manner of a log house and the interior filled with earth or rubble. The heavy crib structure supported the dam's face and the weight of the water.

Timber plank dams were more elegant structures that employed a variety of construction methods utilizing heavy timbers to support a water retaining arrangement of planks.

Very few timber dams are still in use. Timber, in the form of sticks, branches and withes, is the basic material used by beavers, often with the addition of mud or stones.

Steel dams

A steel dam is a type of dam briefly experimented with in around the turn of the 19th-20th Century which uses steel plating (at an angle) and load bearing beams as the structure. Intended as permanent structures, steel dams were an (arguably failed) experiment to determine if a construction technique could be devised that was cheaper than masonry, concrete or earthworks, but sturdier than timber crib dams.

Red Ridge steel dam, b. 1905, Michigan.

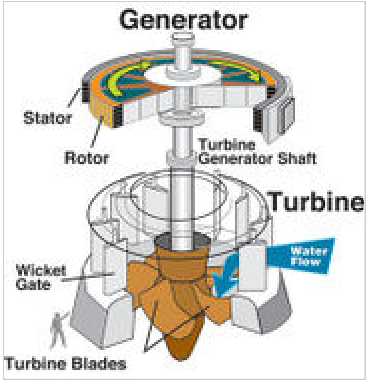
Beaver dams

Beavers create dams primarily out of mud and sticks to flood a particular habitable area. By flooding a parcel of land, beavers can navigate below or near the surface and remain relatively well hidden or protected from predators. The flooded region also allows beavers access to food, especially during the winter.

Construction elements

Power generation plant

As of 2005, hydroelectric power, mostly from dams, supplies some 19% of the world's electricity, and over 63% of renewable energy. Much of this is generated by large dams, although China uses small scale hydro generation on a wide scale and is responsible for about 50% of world use of this type of power.



Hydraulic turbine and electrical generator.

Most hydroelectric power comes from the potential energy of **dammed** water driving a water turbine and generator; to boost the power generation capabilities of a dam, the water may be run through a large pipe called a penstock before the turbine. A variant on this simple model uses pumped storage hydroelectricity to produce electricity to match periods of high and low demand, by moving water between reservoirs at different elevations. At times of low electrical demand, excess generation capacity is used to pump water into the higher reservoir. When there is higher demand, water is released back into the lower reservoir through a turbine.

Spillways

A *spillway* is a section of a dam designed to pass water from the upstream side of a dam to the downstream side. Many spillways have floodgates designed to control the flow through the spillway. Types of spillway include: A *service spillway* or *primary spillway* passes normal flow. An *auxiliary spillway* releases flow in excess of the capacity of the service spillway. An *emergency spillway* is designed for extreme conditions, such as a serious malfunction of the service spillway. A *fuse plug spillway* is a low embankment designed to be over topped and washed away in the event of a large flood. Fusegate elements are independent free-standing block set side by side on the spillway which work without any remote control. They allow to increase the normal pool of the dam without compromising the security of the dam because they are designed to be gradually evacuated for exceptional events. They work as fixed weir most of the time allowing overspilling for the common floods.

Spillway on Llyn Brianne dam, Wales soon after first fill.

The spillway can be gradually eroded by water flow, including cavitation or turbulence of the water flowing over the spillway, leading to its failure. It was the inadequate design of the spillway which led to the 1889 over-topping of the South Fork Dam in Johnstown, Pennsylvania, resulting in the infamous Johnstown Flood (the "great flood of 1889").

Erosion rates are often monitored, and the risk is ordinarily minimized, by shaping the downstream face of the spillway into a curve that minimizes turbulent flow, such as an ogee curve.

Dam creation

Common purposes

Function	Example
Power generation	Hydroelectric power is a major source of electricity in the world. Many countries have rivers with adequate water flow, that can be dammed for power generation purposes. For example, the Itaipu on the Paraná River in South America generates 14 GW and supplied 93% of the energy consumed by Paraguay and 20% of that consumed by Brazil as of 2005.
Stabilize water flow / irrigation	Dams are often used to control and stabilize water *flow*, often for agricultural purposes and irrigation. Others such as the Berg Strait dam can help to stabilize or restore the water *levels* of inland lakes and seas, in this case the Aral Sea.
Flood prevention	Dams such as the Blackwater dam of Webster, New Hampshire and the Delta Works are created with flood control in mind.
Land reclamation	Dams (often called dykes or levees in this context) are used to prevent ingress of water to an area that would otherwise be submerged, allowing its reclamation for human use.
Water diversion	See: diversion dam.

Siting (location)

One of the best places for building a dam is a narrow part of a deep river valley; the valley sides can then act as natural walls. The primary function of the dam's structure is to fill the gap in the natural reservoir line left by the stream channel. The sites are usually those where the gap becomes a minimum for the required storage capacity. The most economical arrangement is often a composite structure such as a masonry dam flanked by earth embankments. The current use of the land to be flooded should be dispensable.

Significant other engineering and engineering geology considerations when building a dam include:

- permeability of the surrounding rock or soil
- earthquake faults
- landslides and slope stability
- peak flood flows
- reservoir silting
- environmental impacts on river fisheries, forests and wildlife (see also fish ladder)
- impacts on human habitations
- compensation for land being flooded as well as population resettlement
- removal of toxic materials and buildings from the proposed reservoir area

Impact assessment

Impact is assessed in several ways: the benefits to human society arising from the dam (agriculture, water, damage prevention and power), the harm or benefits to nature and wildlife (especially fish and rare species), the impact on the geology of an area - whether the change to water flow and levels will increase or decrease stability, and the disruption to human lives (relocation, loss of archeological or cultural matters underwater).

Environmental impact

Wood and garbage accumulated because of a dam

Dams affect many ecological aspects of a river. Rivers depend on the constant disturbance of a certain tolerance. Dams slow the river and this disturbance may damage or destroy this pattern of ecology. Temperature is also another problem that dams create. Rivers tend to have fairly homogeneous temperatures. Reservoirs have layered temperatures, warm on the top and cold on the bottom; in addition often it is water from the colder (lower) layer which is released downstream, and this may have a different dissolved oxygen content than before. Organisms depending upon a regular cycle of temperatures may be unable to adapt; the balance of other fauna (especially plant life and microscopic fauna) may be affected by the change of oxygen content.

Water exiting a turbine usually contains very little suspended sediment, which can lead to scouring of river beds and loss of riverbanks; for example, the daily cyclic flow variation caused by the Glen Canyon Dam was a contributor to sand bar erosion.

Older dams often lack a fish ladder, which keeps many fish from moving up stream to their natural breeding grounds, causing failure of breeding cycles or blocking of migration paths. Even the presence of a fish ladder does not always prevent a reduction in fish reaching the spawning grounds upstream. In some areas, young fish ("smolt") are transported downstream by barge during parts of the year. Turbine and power-plant designs that have a lower impact upon aquatic life are an active area of research.

A large dam can cause the loss of entire ecospheres, including endangered and undiscovered species in the area, and the replacement of the original environment by a new inland lake.

Depending upon the circumstances, a dam can either reduce or increase the net production of greenhouse gases. An **increase** can occur if the reservoir created by the dam itself acts as a source of substantial amounts of potent greenhouse gases (methane and carbon dioxide) due to plant material in flooded areas decaying in an anaerobic environment. According to the World Commission on Dams report, when the reservoir is relatively large and no prior clearing of forest in the flooded area was undertaken, greenhouse gas emissions from the reservoir could be higher than those of a conventional oil-fired thermal generation plant. A **decrease** can occur if the dam is used in place of traditional power generation, since electricity produced from hydroelectric generation does not give rise to any flue gas emissions from fossil fuel combustion (including sulfur dioxide, nitric oxide, carbon monoxide, dust, and mercury from coal).

Large lakes formed behind dams have been indicated as contributing to earthquakes, due to changes in loading and/or the height of the water table.

Human social impact

The impact on human society is also significant. For example, the Three Gorges Dam on the Yangtze River in China, is more than five times the size of the Hoover Dam (U.S.) and will create a reservoir 600 km long, to be used for hydro-power generation. Its construction required the loss of over a million people's homes and their mass relocation, the loss of many valuable archaeological and cultural sites, as well as significant ecological change.

Economics

Construction of a hydroelectric plant requires a long lead-time for site studies, hydrological studies, and environmental impact assessment, and are large scale projects by comparison to traditional power generation based upon fossil fuels. The number of sites that can be economically developed for hydroelectric production is limited; new sites tend to be far from population centers and usually require extensive power transmission lines. Hydroelectric generation can be vulnerable to major changes in the climate, including variation of rainfall, ground and surface water levels, and glacial melt, causing additional expenditure for the extra capacity to ensure sufficient power is available in low water years.

Once completed, if it is well designed and maintained, a hydroelectric power source is usually comparatively cheap and reliable. It has no fuel and low escape risk, and as a renewable energy source it is cheaper than both nuclear and wind power. It is more easily regulated to store water as needed and generate high power levels on demand, compared to wind power.

Dam failure

The reservoir emptying through the failed Teton Dam.

International special sign for works and installations containing dangerous forces

Dam failures are generally catastrophic if the structure is breached or significantly damaged. Routine deformation monitoring of seepage from drains in, and around, larger dams is necessary to anticipate any problems and permit remedial action to be taken before structural failure occurs. Most dams incorporate mechanisms to permit the reservoir to be lowered or even drained in the event of such problems. Another solution can be rock grouting - pressure pumping portland cement slurry into weak fractured rock.

During an armed conflict, a dam is to be considered as an "installation containing dangerous forces" due to the massive impact of a possible destruction on the civilian population and the environment. As such, it is protected by the rules of International Humanitarian Law (IHL) and shall not be made the object of attack if that may cause severe losses among the civilian population. To facilitate the identification, a protective sign consisting of three bright orange circles placed on the same axis is defined by the rules of IHL.

The main causes of dam failure include spillway design error (South Fork Dam), geological instability caused by changes to water levels during filling or poor surveying (Vajont Dam, Malpasset), poor maintenance, especially of outlet pipes (Lawn Lake Dam, Val di Stava Dam collapse), extreme rainfall (Shakidor Dam), and human, computer or design error (Buffalo Creek Flood, Dale Dike Reservoir, Taum Sauk pumped storage plant).

A notable case of deliberate dam failure (prior to the above ruling) was the British Royal Air Force Dambusters raid on

Germany in World War II (codenamed *"Operation Chastise"*), in which three German dams were selected to be breached in order to have an impact on German infrastructure and manufacturing and power capabilities deriving from the Ruhr and Eder rivers. This raid later became the basis for several films.

DVD

2008/9 Schools Wikipedia Selection. Related subjects: Computing hardware and infrastructure; Media

DVD	
Image:Dvd logo.svg	
	
Media type	Optical disc
Capacity	~4.7 GB (single-sided single-layer), ~8.54 GB (single-sided double-layer)
Read mechanism	650 nm laser, 1350 kB/s (1×)
Write mechanism	1350 kB/s (1×)
Usage	Data storage, audio, video, games

DVD (also known as "**Digital Versatile Disc**" or "**Digital Video Disc**" - see Etymology) is a popular optical disc storage media format. Its main uses are video and data storage. Most DVDs are of the same dimensions as compact discs (CDs) but store more than 6 times as much data.

Variations of the term DVD often describe the way data is stored on the discs: DVD-ROM has data which can only be read and not written, DVD-R and DVD+R can be written once and then functions as a DVD-ROM, and DVD-RAM, DVD-RW, or DVD+RW holds data that can be erased and thus re-written multiple times. The wavelength used by standard DVD lasers is 650 nm .

DVD-Video and DVD-Audio discs respectively refer to properly formatted and structured video and audio content. Other types of DVDs, including those with video content, may be referred to as DVD-Data discs. The term "DVD" is commonly misused to refer to high definition optical disc formats in general, such as HD DVD, its official successor, and Blu-ray, its rival.

History

In 1993, two high-density optical storage standards were being developed; one was the MultiMedia Compact Disc, backed by Philips and Sony, and the other was the Super Density disc, supported by Toshiba, Time Warner, Matsushita Electric, Hitachi, Mitsubishi Electric, Pioneer, Thomson, and JVC. IBM's president, Lou Gerstner, acting as a matchmaker, led an effort to unite the two camps behind a single standard, anticipating a repeat of the costly videotape format war between VHS and Betamax in the 1980s.

Philips and Sony abandoned their MultiMedia Compact Disc and fully agreed upon Toshiba's SuperDensity Disc with only one modification, namely changing to EFMPlus modulation. EFMPlus was chosen as it has a great resilience against disc damage such as scratches and fingerprints. EFMPlus, created by Kees Immink, who also designed EFM, is 6% less efficient than the modulation technique originally used by Toshiba, which resulted in a capacity of 4.7 GB as opposed to the original 5 GB. The result was the DVD specification, finalized for the DVD movie player and DVD-ROM computer applications in December 1995.
In May 1997, the DVD Consortium was replaced by the DVD Forum, which is open to all other companies.

Optical disc authoring

Etymology

"DVD" was originally used as an initialism for the unofficial term "digital videodisk". It was reported in 1995, at the time of the specification finalization, that the letters officially stood for "digital versatile disc" (due to non-video applications), however, the text of the press release announcing the specification finalization only refers to the technology as "DVD", making no mention of what (if anything) the letters stood for. A newsgroup FAQ written by Jim Taylor (a prominent figure in the industry) claims that four years later, in 1999, the DVD Forum stated that the format name was simply the three letters "DVD" and did not stand for anything.

The official DVD specification documents have never defined DVD. Usage in the present day varies, with "DVD", "Digital Video Disc", and "Digital Versatile Disc" being the most common.

The DVD Forum website has a section called "DVD Primer" in which the answer to the question, "What does DVD mean?" reads, "The keyword is 'versatile.' Digital Versatile discs provide superb video, audio and data storage and access -- all on one disc."

DVD capacity

	Single layer capacity		Dual/Double layer capacity	
Physical size	**GB**	**GiB**	**GB**	**GiB**
12 cm, single sided	4.7	4.37	8.54	7.95
12 cm, double sided	9.4	8.74	17.08	15.90
8 cm, single sided	1.4	1.30	2.6	2.42
8 cm, double sided	2.8	2.61	5.2	4.84

The 12 cm type is a standard DVD, and the 8 cm variety is known as a mini-DVD. These are the same sizes as a standard CD and a mini-CD.

Note: GB here means gigabyte in the SI sense, i.e. 10^9 (or 1,000,000,000) bytes; GiB is used for gibibyte, equal to 2^{30} (or 1,073,741,824) bytes. For reference, most computer operating systems display file sizes in gibibytes, mebibytes, and kibibytes.

Example: A disc with 8.54 GB capacity is equivalent to: (8.54 × 1,000,000,000) / 1,073,741,824 ≈ 7.95 GiB.

- Optical disc
- Optical disc image
- Optical disc drive
- Optical disc authoring
- Authoring software
- Recording technologies
    - Recording modes
    - Packet writing

Optical media types

- Laserdisc (LD), Video Single Disc (VSD)
- Compact Disc (CD): Red Book, 5.1 Music Disc, SACD, PhotoCD, CD-R, CD-ROM, CD-RW, CD Video (CDV), Video CD (VCD), SVCD, CD+G, CD-Text, CD-ROM XA, CD-i
- MiniDisc (MD) (Hi-MD)
- **DVD**: DVD-R, DVD+R, DVD-R DL, DVD+R DL, DVD-RW, DVD+RW, DVD-RW DL, DVD+RW DL, DVD-RAM, DVD-D
- Ultra Density Optical (UDO)
- Universal Media Disc (UMD)
- HD DVD: HD DVD-R, HD DVD-RW, HD DVD-RAM, HD DVD-ROM
- Blu-ray Disc (BD): BD-R, BD-RE

Standards

- Rainbow Books
- File systems
    - ISO 9660
        - Joliet
        - Rock Ridge
        - El Torito
        - Apple ISO 9660 Extensions
    - Universal Disk Format (ISO 9660) (UDF)
        - Mount Rainier

Each DVD sector contains 2418 bytes of data, 2048 bytes of which are user data.

Capacity Note: There is a small difference in capacity (storage space) between + and - DL DVD formats. For example, the 12 cm single sided disc has capacities:

Size comparison: A 12 cm Sony DVD+RW and a 19 cm pencil.

Disc Type	Sectors	bytes	GB	GiB
DVD-R SL	2,298,496	4,707,319,808	4.7	4.384
DVD+R SL	2,295,104	4,700,372,992	4.7	4.378
DVD-R DL	4,171,712	8,543,666,176	8.5	7.957
DVD+R DL	4,173,824	8,547,991,552	8.5	7.961

Capacity nomenclature

The five basic types of DVD are referred to by their approximate capacity in gigabytes.

DVD type	Name
Single sided, single layer	DVD-5
Single sided, dual layer	DVD-9
Double sided, single layer	DVD-10
Double sided, dual layer on one side, single on other	DVD-14
Double sided, dual layer on both sides	DVD-18

Technology

DVD uses 650 nm wavelength laser diode light as opposed to 780 nm for CD. This permits a smaller spot on the media surface (1.32 µm for DVD versus 2.11 µm for CD).

Internal mechanism of a DVD-ROM Drive

Writing speeds for DVD were 1×, that is 1350 kB/s (1318 KiB/s), in the first drives and media models. More recent models at 18× or 20× have 18 or 20 times that speed. Note that for CD drives, 1× means 153.6 kB/s (150 KiB/s), 9 times slower. DVD FAQ

Speed

Drive speed	Data rate		Write time for Single Layer DVD
1X	10.55 Mbit/s	1.32 MB/s	61 min.
2X	21.09 Mbit/s	2.64 MB/s	30 min.
4X	42.19 Mbit/s	5.27 MB/s	15 min.
8X	84.38 Mbit/s	10.55 MB/s	8 min.
16X	168.75 Mbit/s	21.09 MB/s	4 min.

DVD recordable and rewritable

HP initially developed recordable DVD media from the need to store data for back-up and transport.

DVD recordables are now also used for consumer audio and video recording. Three formats were developed: DVD- R/ RW (minus/dash), DVD+ R/ RW (plus), DVD-RAM.

Dual layer recording

Dual Layer recording allows DVD-R and DVD+R discs to store significantly more data, up to 8.5 gigabytes per side, per disc, compared with 4.7 gigabytes for single-layer discs. DVD-R DL was developed for the DVD Forum by Pioneer Corporation, DVD+R DL was developed for the DVD+RW Alliance by Philips and Mitsubishi Kagaku Media (MKM).

A Dual Layer disc differs from its usual DVD counterpart by employing a second physical layer within the disc itself. The drive with Dual Layer capability accesses the second layer by shining the laser through the first semi-transparent layer. The layer change mechanism in some DVD players can show a noticeable pause, as long as two seconds by some accounts. This caused more than a few viewers to worry that their dual layer discs were damaged or defective, with the end result that studios began listing a standard message explaining the dual layer pausing effect on all dual layer disc packaging.

DVD recordable discs supporting this technology are backward compatible with some existing DVD players and DVD-ROM drives. Many current DVD recorders support dual-layer technology, and the price is now comparable to that of single-layer drives, though the blank media remains more expensive. The recording speeds reached by dual-layer media are still well below those of single-layer media.

DVD-Video

DVD-Video is a standard for storing video content on DVD media. In the U.S., weekly DVD-Video rentals first out-numbered weekly VHS cassette rentals in June 2003, illustrating the rapid adoption rate of the technology in the marketplace.

Though many resolutions and formats are supported, most consumer DVD-Video discs use either 4:3 or anamorphic 16:9 aspect ratio MPEG-2 video, stored at a resolution of 720×480 (NTSC) or 720×576 (PAL) at 29.97 or 25 FPS. Audio is commonly stored using the Dolby Digital (AC-3) or Digital Theatre System (DTS) formats, ranging from 16-bits/48 kHz to 24-bits/96 kHz format with monaural to 7.1 channel " Surround Sound" presentation, and/or MPEG-1 Layer 2. Although the specifications for video and audio requirements vary by global region and television system, many DVD players support all possible formats. DVD-Video also supports features like menus, selectable subtitles, multiple camera angles, and multiple audio tracks.

DVD-Audio

DVD-Audio is a format for delivering high-fidelity audio content on a DVD. It offers many channel configuration options (from mono to 7.1 surround sound) at various sampling frequencies (up to 24-bits/192 kHz versus CDDAs 16-bits/44.1 kHz). Compared with the CD format, the much higher capacity DVD format enables the inclusion of considerably more music (with respect to total running time and quantity of songs) and/or far higher audio quality (reflected by higher linear sampling rates and higher vertical bit-rates, and/or additional channels for spatial sound reproduction).

Despite DVD-Audio's superior technical specifications, there is debate as to whether the resulting audio enhancements are distinguishable in typical listening environments. DVD-Audio currently forms a niche market, probably due to the very sort of format war with rival standard SACD that DVD-Video avoided.

Security

DVD-Audio discs employ a robust copy prevention mechanism, called Content Protection for Prerecorded Media (CPPM) developed by the 4C group (IBM, Intel, Matsushita, and Toshiba).

To date, CPPM has not been "broken" in the sense that DVD-Video's CSS has been broken, but ways to circumvent it have been developed. By modifying commercial DVD(-Audio) playback software to write the decrypted and decoded audio streams to the hard disk, users can, essentially, extract content from

DVD-Audio discs much in the same way they can from DVD-Video discs.

Successors

There are several possible successors to DVD being developed by different consortia. Sony/ Panasonic's Blu-ray Disc (BD) as a rival to DVD forum's HD DVD, the "Official" successor designed by Toshiba. Both began to gain traction in 2007, and next-generation technologies such as Maxell's Holographic Versatile Disc (HVD) and 3D optical data storage are being actively developed.

On November 19, 2003, the DVD Forum decided by a vote of eight to six that HD DVD will be its official HD successor to DVD. In spite of this, both BD and HD DVD have already severely hampered the adoption of any successor to DVD through a lack of the very cooperation that fostered DVD's success.

Electrical engineering

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Electrical Engineers design complex power systems...

... and complex electronic circuits.

Electrical engineering, sometimes referred to as **electrical and electronic engineering**, is a field of engineering that deals with the study and application of electricity, electronics and electromagnetism. The field first became an identifiable occupation in the late nineteenth century after commercialization of the electric telegraph and electrical power supply. It now covers a range of subtopics including power, electronics, control systems, signal processing and telecommunications.

Electrical engineering may or may not encompass electronic engineering. Where a distinction is made, usually outside of the United States, electrical engineering is considered to deal with the problems associated with large-scale electrical systems such as power transmission and motor control, whereas electronic engineering deals with the study of small-scale electronic systems including computers and integrated circuits. Alternatively, electrical engineers are usually concerned with using electricity to transmit energy, while electronic engineers are concerned with using electricity to transmit information.

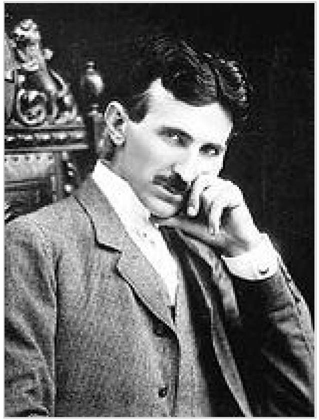
History

Electricity has been a subject of scientific interest since at least the early 17th century. The first electrical engineer was probably William Gilbert who designed the versorium: a device that detected the presence of statically charged objects. He was also the first to draw a clear distinction between magnetism and static electricity and is credited with establishing the term electricity. In 1775 Alessandro Volta's scientific experimentations devised the electrophorus, a device that produced a static electric charge, and by 1800 Volta developed the voltaic pile, a forerunner of the electric battery.

However, it was not until the 19th century that research into the subject started to intensify. Notable developments in this century include the work of Georg Ohm, who in 1827 quantified the relationship between the electric current and potential difference in a conductor, Michael Faraday, the discoverer of electromagnetic induction in 1831, and James Clerk Maxwell, who in 1873 published a unified theory of electricity and magnetism in his treatise *Electricity and Magnetism*.

During these years, the study of electricity was largely considered to be a subfield of physics. It was not until the late 19th century that universities started to offer degrees in electrical engineering. The Darmstadt University of Technology founded the first chair and the first faculty of electrical engineering worldwide in 1882. In 1883 Darmstadt University of Technology and Cornell University introduced the world's first courses of study in electrical engineering, and in 1885 the University College London founded the first chair of electrical engineering in the United Kingdom. The University of Missouri subsequently established the first department of electrical engineering in the United States in 1886.


Thomas Edison built the world's first large-scale electrical supply network


Nikola Tesla made long-distance electrical transmission networks possible.

During this period, the work concerning electrical engineering increased dramatically. In 1882, Edison switched on the world's first large-scale electrical supply network that provided 110 volts direct current to fifty-nine customers in lower Manhattan. In 1887, Nikola Tesla filed a number of patents related to a competing form of power distribution known as alternating current. In the following years a bitter rivalry between Tesla and Edison, known as the " War of Currents", took place over the preferred method of distribution. AC eventually replaced DC for generation and power distribution, enormously extending the range and improving the safety and efficiency of power distribution.

The efforts of the two did much to further electrical engineering—Tesla's work on induction motors and polyphase systems influenced the field for years to come, while Edison's work on telegraphy and his development of the stock ticker proved lucrative for his company, which ultimately became General Electric. However, by the end of the 19th century, other key figures in the progress of electrical engineering were beginning to emerge.

Modern developments

Emergence of radio and electronics

During the development of radio, many scientists and inventors contributed to radio technology and electronics. In his classic UHF experiments of 1888, Heinrich Hertz transmitted (via a spark-gap transmitter) and detected radio waves using electrical equipment. In 1895, Nikola Tesla was able to detect signals from the transmissions of his New York lab at West Point (a distance of 80.4 km / 49.95 miles). In 1897, Karl Ferdinand Braun introduced the cathode ray tube as part of an oscilloscope, a crucial enabling technology for electronic television. John Fleming invented the first radio tube, the diode, in 1904. Two years later, Robert von Lieben and Lee De Forest independently developed the amplifier tube, called the triode. In 1895, Guglielmo Marconi furthered the art of hertzian wireless methods. Early on, he sent wireless signals over a distance of one and a half miles. In December 1901, he sent wireless waves that were not affected by the curvature of the Earth. Marconi later transmitted the wireless signals across the Atlantic between Poldhu, Cornwall, and St. John's, Newfoundland, a distance of 2100 miles. In 1920 Albert Hull developed the magnetron which would eventually lead to the development of the microwave oven in 1946 by Percy Spencer. In 1934 the British military began to make strides towards radar (which also uses the magnetron) under the direction of Dr Wimperis, culminating in the operation of the first radar station at Bawdsey in August 1936.

In 1941 Konrad Zuse presented the Z3, the world's first fully functional and programmable computer. In 1946 the ENIAC (Electronic Numerical Integrator and Computer) of John Presper Eckert and John Mauchly followed, beginning the computing era. The arithmetic performance of these machines allowed engineers to develop completely new technologies and achieve new objectives, including the Apollo missions and the NASA moon landing.

The invention of the transistor in 1947 by William B. Shockley, John Bardeen and Walter Brattain opened the door for more compact devices and led to the development of the integrated circuit in 1958 by Jack Kilby and independently in 1959 by Robert Noyce. In 1968 Marcian Hoff invented the first microprocessor at Intel and thus ignited the development of the personal computer. The first realization of the microprocessor was the Intel 4004, a 4-bit processor developed in 1971, but only in 1973 did the Intel 8080, an 8-bit processor, make the building of the first personal computer, the Altair 8800, possible.

Education

Electrical engineers typically possess an academic degree with a major in electrical engineering. The length of study for such a degree is usually four or five years and the completed degree may be designated as a Bachelor of Engineering, Bachelor of Science, Bachelor of Technology or Bachelor of Applied Science depending upon the university. The degree generally includes units covering physics, mathematics, computer science, project management and specific topics in electrical engineering. Initially such topics cover most, if not all, of the sub-disciplines of electrical engineering. Students then choose to specialize in one or more sub-disciplines towards the end of the degree.

Some electrical engineers also choose to pursue a postgraduate degree such as a Master of Engineering/ Master of Science (MEng/MSc), a Master of Engineering Management, a Doctor of Philosophy (PhD) in Engineering, an Engineering Doctorate (EngD), or an Engineer's degree. The Master and Engineer's degree may consist of either research, coursework or a mixture of the two. The Doctor of Philosophy and Engineering Doctorate degrees consist of a significant research component and are often viewed as the entry point to academia. In the United Kingdom and various other European countries, the Master of Engineering is often considered an undergraduate degree of slightly longer duration than the Bachelor of Engineering.

Practicing engineers

In most countries, a Bachelor's degree in engineering represents the first step towards professional certification and the degree program itself is certified by a professional body. After completing a certified degree program the engineer must satisfy a range of requirements (including work experience requirements) before being certified. Once certified the engineer is designated the title of Professional Engineer (in the United States, Canada and South Africa), Chartered Engineer (in India, the United Kingdom, Ireland and Zimbabwe), Chartered Professional Engineer (in Australia and New Zealand) or European Engineer (in much of the European Union).

The advantages of certification vary depending upon location. For example, in the United States and Canada "only a licensed engineer may seal engineering work for public and private clients". This requirement is enforced by state and provincial legislation such as Quebec's Engineers Act. In other countries, such as Australia, no such legislation exists. Practically all certifying bodies maintain a code of ethics that they expect all members to abide by or risk expulsion. In this way these organizations play an important role in maintaining ethical standards for the profession. Even in jurisdictions where certification has little or no legal bearing on work, engineers are subject to contract law. In cases where an engineer's work fails he or she may be subject to the tort of negligence and, in extreme cases, the charge of criminal negligence. An engineer's work must also comply with numerous other rules and regulations such as building codes and legislation pertaining to environmental law.

Professional bodies of note for electrical engineers include the Institute of Electrical and Electronics Engineers (IEEE) and the Institution of Engineering and Technology (IET) (which was formed by the merging of the Institution of Electrical Engineers (IEE) and the Institution of Incorporated Engineers (IIE). The IEEE claims to produce 30% of the world's literature in electrical engineering, has over 360,000 members worldwide and holds over 3,000 conferences annually. The IET publishes 21 journals, has a worldwide membership of over 150,000, and claims to be the largest professional engineering society in Europe. Obsolescence of technical skills is a serious concern for electrical engineers. Membership and participation in technical societies, regular reviews of periodicals in the field and a habit of continued learning are therefore essential to maintaining proficiency.

In countries such as Australia, Canada and the United States electrical engineers make up around 0.25% of the labor force (see note). Outside of these countries, it is difficult to gauge the demographics of the profession due to less meticulous reporting on labour statistics. However, in terms of electrical engineering graduates per-capita, electrical engineering graduates would probably be most numerous in countries such as Taiwan, Japan, India and South Korea.

Tools and work

From the Global Positioning System to electric power generation, electrical engineers have contributed to the development of a wide range of technologies. They design, develop, test and supervise the deployment of electrical systems and electronic devices. For example, they may work on the design of telecommunication systems, the operation of electric power stations, the lighting and wiring of buildings, the design of household appliances or the electrical control of industrial machinery.

Satellite communications is one of many projects an electrical engineer might work on

Fundamental to the discipline are the sciences of physics and mathematics as these help to obtain both a qualitative and quantitative description of how such systems will work. Today most engineering work involves the use of computers and it is commonplace to use computer-aided design programs when designing electrical systems. Nevertheless, the ability to sketch ideas is still invaluable for quickly communicating with others.

Although most electrical engineers will understand basic circuit theory (that is the interactions of elements such as resistors, capacitors, diodes, transistors and inductors in a circuit), the theories employed by engineers generally depend upon the work they do. For example, quantum mechanics and solid state physics might be relevant to an engineer working on VLSI (the design of integrated circuits), but are largely irrelevant to engineers working with macroscopic electrical systems. Even circuit theory may not be relevant to a person designing telecommunication systems that use off-the-shelf components. Perhaps the most important technical skills for electrical engineers are reflected in university programs, which emphasize strong numerical skills, computer literacy and the ability to understand the technical language and concepts that relate to electrical engineering.

For many engineers, technical work accounts for only a fraction of the work they do. A lot of time may also be spent on tasks such as discussing proposals with clients, preparing budgets and determining project schedules. Many senior engineers manage a team of technicians or other engineers and for this reason project management skills are important. Most engineering projects involve some form of documentation and strong written communication skills are therefore very important.

The workplaces of electrical engineers are just as varied as the types of work they do. Electrical engineers may be found in the pristine lab environment of a fabrication plant, the offices of a consulting firm or on site at a mine. During their working life, electrical engineers may find themselves supervising a wide range of individuals including scientists, electricians, computer programmers and other engineers.

Sub-disciplines

Electrical engineering has many sub-disciplines, the most popular of which are listed below. Although there are electrical engineers who focus exclusively on one of these sub-disciplines, many deal with a combination of them. Sometimes certain fields, such as electronic engineering and computer engineering, are considered separate disciplines in their own right.

Power

Power engineering deals with the generation, transmission and distribution of electricity as well as the design of a range of related devices. These include transformers, electric generators, electric motors, high voltage engineering and power electronics. In many regions of the world, governments maintain an electrical network called a power grid that connects a variety of generators together with users of their energy. Users purchase electrical energy from the grid, avoiding the costly exercise of having to generate their own. Power engineers may work on the design and maintenance of the power grid as well as the power systems that connect to it. Such systems are called *on-grid* power systems and may supply the grid with additional power, draw power from the grid or do both. Power engineers may also work on systems that do not connect to the grid, called *off-grid* power systems, which in some cases are preferable to on-grid systems. The future includes Satellite controlled power systems, with feedback in real time to prevent power surges and prevent blackouts.

Control

Control systems play a critical role in space flight

Control engineering focuses on the modeling of a diverse range of dynamic systems and the design of controllers that will cause these systems to behave in the desired manner. To implement such controllers electrical engineers may use electrical circuits, digital signal processors, microcontrollers and PLCs (Programmable Logic Controllers). Control engineering has a wide range of applications from the flight and propulsion systems of commercial airliners to the cruise control present in many modern automobiles. It also plays an important role in industrial automation.

Control engineers often utilize feedback when designing control systems. For example, in an automobile with cruise control the vehicle's speed is continuously monitored and fed back to the system which adjusts the motor's power output accordingly. Where there is regular feedback, control theory can be used to determine how the system responds to such feedback.

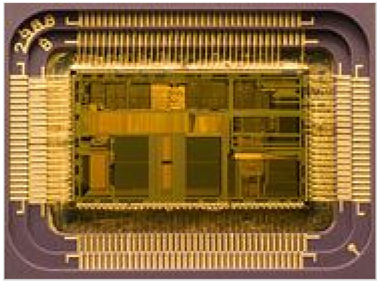
Electronics

Electronic engineering involves the design and testing of electronic circuits that use the properties of components such as resistors, capacitors, inductors, diodes and transistors to achieve a particular functionality. The tuned circuit, which allows the user of a radio to filter out all but a single station, is just one example of such a circuit. Another example (of a pneumatic signal conditioner) is shown in the adjacent photograph.

Prior to the second world war, the subject was commonly known as *radio engineering* and basically was restricted to aspects of communications and radar, commercial radio and early television. Later, in post war years, as consumer devices began to be developed, the field grew to include modern television, audio systems, computers and microprocessors. In the mid to late 1950s, the term *radio engineering* gradually gave way to the name *electronic engineering*.

Before the invention of the integrated circuit in 1959, electronic circuits were constructed from discrete components that could be manipulated by humans. These discrete circuits consumed much space and power and were limited in speed, although they are still common in some applications. By contrast, integrated circuits packed a large number—often millions—of tiny electrical components, mainly transistors, into a small chip around the size of a coin. This allowed for the powerful computers and other electronic devices we see today.

Microelectronics

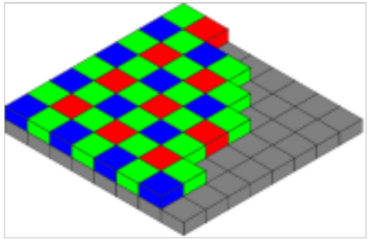


Microprocessor

Microelectronics engineering deals with the design of very small electronic circuit components for use in an integrated circuit or sometimes for use on their own as a general electronic component. The most common microelectronic components are semiconductor transistors, although all main electronic components (resistors, capacitors, inductors) can be created at a microscopic level.

Microelectronic components are created by chemically fabricating wafers of semiconductors such as silicon (at higher frequencies, compound semiconductors like gallium arsenide and indium phosphide) to obtain the desired transport of electronic charge and control of current. The field of microelectronics involves a significant amount of chemistry and material science and requires the electronic engineer working in the field to have a very good working knowledge of the effects of quantum mechanics.

Signal processing



A Bayer filter on a CCD requires signal processing to get a red, green, and blue value at each pixel

Signal processing deals with the analysis and manipulations of signals. Signals can be either analog, in which case the signal varies continuously according to the information, or digital, in which case the signal varies according to a series of discrete values representing the information. For analog signals, signal processing may involve the amplification and filtering of audio signals for audio equipment or the modulation and demodulation of signals for telecommunications. For digital signals, signal processing may involve the compression, error detection and error correction of digitally sampled signals.

Telecommunications

Telecommunications engineering focuses on the transmission of information across a channel such as a coax cable, optical fibre or free space. Transmissions across free space require information to be encoded in a carrier wave in order to shift the information to a carrier frequency suitable for transmission, this is known as modulation. Popular analog modulation techniques include amplitude modulation and frequency modulation. The choice of modulation affects the cost and performance of a system and these two factors must be balanced carefully by the engineer.

Once the transmission characteristics of a system are determined, telecommunication engineers design the transmitters and receivers needed for such systems. These two are sometimes combined to form a two-way communication device known as a transceiver. A key consideration in the design of transmitters is their power consumption as this is closely related to their signal strength. If the signal strength of a transmitter is insufficient the signal's information will be corrupted by noise.

Instrumentation engineering

Instrumentation engineering deals with the design of devices to measure physical quantities such as pressure, flow and temperature. The design of such instrumentation requires a good understanding of physics that often extends beyond electromagnetic theory. For example, radar guns use the Doppler effect to measure the speed of oncoming vehicles. Similarly, thermocouples use the Peltier-Seebeck effect to measure the temperature difference between two points.

Often instrumentation is not used by itself, but instead as the sensors of larger electrical systems. For example, a thermocouple might be used to help ensure a furnace's temperature remains constant. For this reason, instrumentation engineering is often viewed as the counterpart of control engineering.

Computers

Computer engineering deals with the design of computers and computer systems. This may involve the design of new hardware, the design of PDAs or the use of computers to control an industrial plant. Computer engineers may also work on a system's software. However, the design of complex software systems is often the domain of software engineering, which is usually considered a separate discipline. Desktop computers represent a tiny fraction of the devices a computer engineer might work on, as computer-like architectures are now found in a range of devices including video game consoles and DVD players.

Related disciplines

Mechatronics is an engineering discipline which deals with the convergence of electrical and mechanical systems. Such combined systems are known as electromechanical systems and have widespread adoption. Examples include automated manufacturing systems, heating, ventilation and air-conditioning systems and various subsystems of aircraft and automobiles.

The term *mechatronics* is typically used to refer to macroscopic systems but futurists have predicted the emergence of very small electromechanical devices. Already such small devices, known as micro electromechanical systems (MEMS), are used in automobiles to tell airbags when to deploy, in digital projectors to create sharper images and in inkjet printers to create nozzles for high definition printing. In the future it is hoped the devices will help build tiny implantable medical devices and improve optical communication.

Biomedical engineering is another related discipline, concerned with the design of medical equipment. This includes fixed equipment such as ventilators, MRI scanners and electrocardiograph monitors as well as mobile equipment such as cochlear implants, artificial pacemakers and artificial hearts.

Electricity

2008/9 Schools Wikipedia Selection. Related subjects: Electricity and Electronics; Physics


Lightning is one of the most prominent effects of electricity

Electricity (from New Latin *ēlectricus*, "amber-like") is a general term that encompasses a variety of phenomena resulting from the presence and flow of electric charge. These include many easily recognisable phenomena such as lightning and static electricity, but in addition, less familiar concepts such as the electromagnetic field and electromagnetic induction.

In general usage, the word 'electricity' is adequate to refer to a number of physical effects. However, in scientific usage, the term is vague, and these related, but distinct, concepts are better identified by more precise terms:

- **Electric charge** – a property of some subatomic particles, which determines their electromagnetic interactions. Electrically charged matter is influenced by, and produces, electromagnetic fields.
- **Electric current** – a movement or flow of electrically charged particles, typically measured in amperes.
- **Electric field** – an influence produced by an electric charge on other charges in its vicinity.
- **Electric potential** – the capacity of an electric field to do work, typically measured in volts.
- **Electromagnetism** – a fundamental interaction between the electric field and the presence and motion of electric charge.

Electricity has been studied since antiquity, though scientific advances were not forthcoming until the seventeenth and eighteenth centuries. It would remain however until the late nineteenth century that engineers were able to put electricity to industrial and residential use, a time which witnessed a rapid expansion in the development of electrical technology. Electricity's extraordinary versatility as a source of energy means it can be put to an almost limitless set of applications which include transport, heating, lighting, communications, and computation. The backbone of modern industrial society is, and for the foreseeable future can be expected to remain, the use of electrical power.

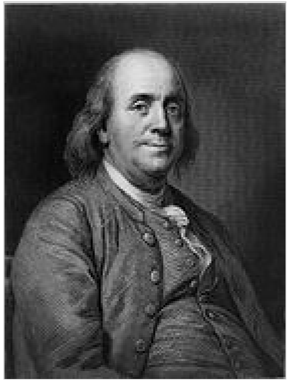
History

Thales, the earliest researcher into electricity

That certain objects such as rods of amber could be rubbed with cat's fur and attract light objects like feathers was known to ancient cultures around the Mediterranean. Thales of Miletos conducted a series of experiments into static electricity around 600 BC, from which he believed that friction rendered amber magnetic, in contrast to minerals such as magnetite, which needed no rubbing. Thales was incorrect in believing the attraction was due to a magnetic effect, but later science would prove a link between magnetism and electricity.

A controversial claim is made that the Parthians and Mesopotamians had some knowledge of electroplating, based on the 1936 discovery of the Baghdad Battery, which resembles a galvanic cell, though this claims lacks evidence supporting the exact nature of the artefact, and whether it was electrical in nature.

Several ancient writers, such as Pliny the Elder and Scribonius Largus, attested to the numbing effect of electric shocks delivered by catfish and torpedo rays, and knew that such shocks could travel along conducting objects. Patients suffering from ailments such as gout or headache were directed to touch electric fish in the hope that the powerful jolt might cure them.



Benjamin Franklin conducted extensive research on electricity in the 18th century

Electricity would remain little more than an intellectual curiosity for over two millennia until 1600, when the English physician William Gilbert made a careful study of electricity and magnetism, distinguishing the lodestone effect from static electricity produced by rubbing amber. He coined the New Latin word *electricus* ("of amber" or "like amber", from *ηλεκτρον* [*elektron*], the Greek word for "amber") to refer to the property of attracting small objects after being rubbed. This association gave rise to the English words "electric" and "electricity", which made their first appearance in print in Sir Thomas Browne's *Pseudodoxia Epidemica* of 1646.

Further work was conducted by Otto von Guericke, Robert Boyle, Stephen Gray and C. F. du Fay. In the 18th century, Benjamin Franklin conducted extensive research in electricity, selling his possessions to fund his work. In June 1752 he is reputed to have attached a metal key to the bottom of a dampened kite string and flown the kite in a storm-threatened sky. He observed a succession of sparks jumping from the key to the back of his hand, showing that lightning was indeed electrical in nature.

In 1791 Luigi Galvani published his discovery of bioelectricity, demonstrating that electricity was the medium by which nerve cells passed signals to the muscles. Alessandro Volta's battery, or voltaic pile, of 1800, made from alternating layers of zinc and copper, provided scientists with a more reliable source of electrical energy than the electrostatic machines previously used. André-Marie Ampère discovered the relationship between electricity and magnetism in 1820; Michael Faraday invented the electric motor in 1821, and Georg Ohm mathematically analysed the electrical circuit in 1827.

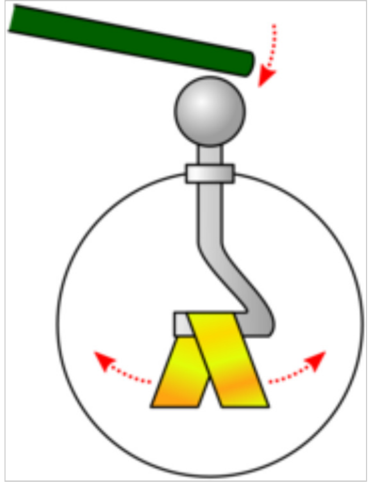
While it had been the early nineteenth century that had seen rapid progress in electrical science, the late nineteenth century would see the greatest progress in electrical engineering. Through such people as Nikola Tesla, Thomas Edison, George Westinghouse, Werner von Siemens, Alexander Graham Bell and Lord Kelvin, electricity was turned from a scientific curiosity into an essential tool for modern life, becoming a driving force for the Second Industrial Revolution.

Concepts

Electric charge

Electric charge is a property of certain subatomic particles, which gives rise to and interacts with, the electromagnetic force, one of the four fundamental forces of nature. Charge originates in the atom, in which its most familiar carriers are the electron and proton. It is a conserved quantity, that is, the net charge within an isolated system will always remain constant regardless of any changes taking place within that system. Within the system, charge may be transferred between bodies, either by direct contact, or by passing along a conducting material, such as a wire. The informal term static electricity refers to the net presence (or 'imbalance') of charge on a body, usually caused when dissimilar materials are rubbed together, transferring charge from one to the other.

The presence of charge gives rise to the electromagnetic force: charges exert a force on each other, an effect that was known, though not understood, in antiquity. A lightweight ball suspended from a string can be charged by touching it with a glass rod that has itself been charged by rubbing with a cloth. If a similar ball is charged by the same glass rod, it is found to repel the first: the charge acts to force the two balls apart. Two balls that are charged with an rubbed amber rod also repel each other. However, if one ball is charged by the glass rod, and the other by an amber rod, the two balls are found to attract each other. These phenomena were investigated by Charles-Augustin de Coulomb in the late eighteenth century, who deduced that charge manifests itself in two opposing forms, leading to the well-known axiom: *like-charged objects repel and opposite-charged objects attract*.



Charge on a gold-leaf electroscope causes the leaves to visibly repel each other

The force acts on the charged particles themselves, hence charge has a tendency to spread itself as evenly as possible over a conducting surface. The magnitude of the electromagnetic force, whether attractive or repulsive, is given by Coulomb's Law, which relates the force to the product of the charges and has an inverse square relation to the distance between them. The electromagnetic force is very strong, second only in strength to the strong interaction, but unlike that force it operates over all distances. In comparison with the much weaker gravitational force, the electromagnetic force pushing two electrons apart is 10^{42} times that of the gravitational attraction pulling them together.

The charge on electrons and protons is opposite in sign, hence an amount of charge may be expressed as being either negative or positive. By convention, the charge carried by electrons is deemed negative, and that by protons positive, a custom that originated with the work of Benjamin Franklin. The amount of charge is usually given the symbol Q and expressed in coulombs; each electron carries the same charge of approximately -1.6022×10^{-19} coulomb. The proton has a charge that is equal and opposite, and thus $+1.6022\times10^{-19}$ coulomb. Charge is possessed not just by matter, but also by antimatter, each antiparticle bearing an equal and opposite charge to its corresponding particle.

Charge can be measured by a number of means, an early instrument being the gold-leaf electroscope, which although still in use for classroom demonstrations, has been superseded by the electronic electrometer.

Electric current

The movement of electric charge is known as an electric current, the intensity of which is usually measured in amperes. Current can consist of any moving charged particles; most commonly these are electrons, but any charge in motion constitutes a current.

By historical convention, a positive current is defined as having the same direction of flow as any positive charge it contains, or to flow from the most positive part of a circuit to the most negative part. Current defined in this manner is called conventional current. The motion of negatively-charged electrons around an electric circuit, one of the most familiar forms of current, is thus deemed positive in the *opposite* direction to that of the electrons. However, depending on the conditions, an electric current can consist of a flow of charged particles in either direction, or even in both directions at once. The positive-to-negative convention is widely used to simplify this situation. If another definition is used—for example, "electron current"—it needs to be explicitly stated.

An electric arc provides an energetic demonstration of electric current

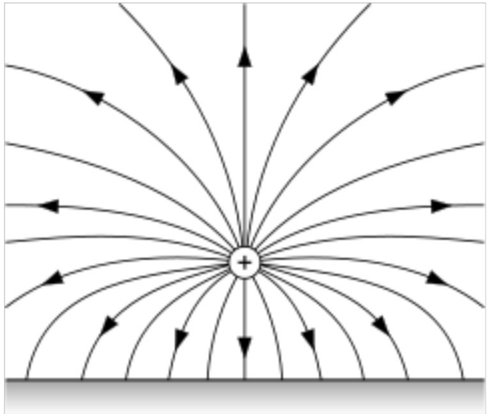
The process by which electric current passes through a material is termed electrical conduction, and its nature varies with that of the charged particles and the material through which they are travelling. Examples of electric currents include metallic conduction, where electrons flow through a conductor such as metal, and electrolysis, where ions (charged atoms) flow through liquids. While the particles themselves can move quite slowly, sometimes with a average drift velocity only fractions of a millimetre per second, the electric field that drives them itself propagates at close to the speed of light, enabling electrical signals to pass rapidly along wires.

Current causes several observable effects, which historically were the means of recognising its presence. That water could be decomposed by the current from a voltaic pile was discovered by Nicholson and Carlisle in 1800, a process now known as electrolysis. Their work was greatly expanded upon by Michael Faraday in 1833. Current through a resistance causes localised heating, an effect James Joule studied mathematically in 1840. One of the most important discoveries relating to current was made accidentally by Hans Christian Ørsted in 1820, when, while preparing a lecture, he witnessed the current in a wire disturbing the needle of a magnetic compass. He had discovered electromagnetism, a fundamental interaction between electricity and magnetics.

In engineering or household applications, current is often described as being either direct current (DC) or alternating current (AC). These terms refer to how the current varies in time. Direct current, as produced by example from a battery and required by most electronic devices, is a unidirectional flow from the positive part of a circuit to the negative. If, as is most common, this flow is carried by electrons, they will be travelling in the opposite direction. Alternating current is any current that reverses direction repeatedly; almost always this takes the form of a sinusoidal wave. Alternating current thus pulses back and forth within a conductor without the charge moving any net distance over time. The time-averaged value of an alternating current is zero, but it delivers energy in first one direction, and then the reverse. Alternating current is affected by electrical properties that are not observed under steady-state direct current, such as inductance and capacitance. These properties however can become important when circuitry is subjected to transients, such as when first energised.

Electric field

The concept of the electric field was introduced by Michael Faraday. An electric field is created by a charged body in the space that surrounds it, and results in a force exerted on any other charges placed within the field. The electric field acts between two charges in a similar manner to the way that the gravitational field acts between two masses, and like it, extends towards infinity and shows an inverse square relationship with distance. However, there is an important difference. Gravity always acts in attraction, drawing two masses together, while the electric field can result in either attraction or repulsion. Since large bodies such as planets generally carry no net charge, the electric field at a distance is usually zero. Thus gravity is the dominant force at distance in the universe, despite being much the weaker.



Field lines emanating from a positive charge above a plane conductor

An electric field generally varies in space, and its strength at any one point is defined as the force (per unit charge) that *would* be felt by a stationary, negligible charge *if* placed at that point. The conceptual charge, termed a *test charge*, must be vanishingly small to prevent its own electric field disturbing the main field and must also be stationary to prevent the effect of magnetic fields. As the electric field is defined in terms of force, and force is a vector, so it follows that an electric field is also a vector, having both magnitude and direction. Specifically, it is a vector field.

The study of electric fields created by stationary charges is called electrostatics. The field may be visualised by a set of imaginary lines whose direction at any point is the same as that of the field. This concept was introduced by Faraday, whose term ' lines of force' still sometimes sees use. The field lines are the paths that a point positive charge would seek to make as it was forced to move within the field; they are however an imaginary concept with no physical existence, and the field permeates all the intervening space between the lines. Field lines emanating from stationary charges have several key properties: first, that they originate at positive charges and terminate at negative charges; second, that they must enter any good conductor at right angles, and third, that they may never cross nor close in on themselves.

The principals of electrostatics are important when designing items of high-voltage equipment. There is a finite limit to the electric field strength that may withstood by any medium. Beyond this point, electrical breakdown occurs and an electrical arc causes flashover between the charged parts. Air, for example, tends to arc at electric field strengths which exceed 30 kV per centimetre across small gaps. Over larger gaps, its breakdown strength is weaker, perhaps 1 kV per centimetre. The most visible natural occurrence of this is lightning, caused when charge becomes separated in the clouds by rising columns of air, and raises the electric field in the air to greater than it can withstand. The voltage of a large lightning cloud may be as high as 100 MV and have discharge energies as great as 250 kWh.

The field strength is greatly affected by nearby conducting objects, and it is particularly intense when it is forced to curve around sharply pointed objects. This principal is exploited in the lightning conductor, the sharp spike of which acts to encourage the lightning stroke to develop there, rather than to the building it serves to protect.

Electric potential

A pair of AA cells. The + sign indicates the terminal from which conventional current moves.

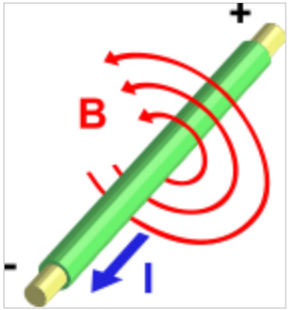
The concept of electric potential is closely linked to that of the electric field. A small charge placed within an electric field experiences a force, and to have brought that charge to that point against the force requires work. The electric potential at any point is defined as the energy required to bring a unit test charge from an infinite distance slowly to that point. It is usually measured in volts, and one volt is the potential for which one joule of work must be expended to bring a charge of one coulomb from infinity. This definition of potential, while formal, has little practical application, and a more useful concept is that of electric potential difference, and is the energy required to move a unit charge between two specified points. An electric field has the special property that it is *conservative*, which means that the path taken by the test charge is irrelevant: all paths between two specified points expend the same energy, and thus a unique value for potential difference may be stated. The volt is so strongly identified as the unit of choice for measurement and description of electric potential difference that the term voltage sees greater everyday usage.

For practical purposes, it is useful to define a common reference point to which potentials may be expressed and compared. While this could be at infinity, a much more useful reference is the Earth itself, which is assumed to be at the same potential everywhere. This reference point naturally takes the name earth or ground. Earth is assumed to be an infinite source of equal amounts of positive and negative charge, and is therefore electrically uncharged – and unchargeable.

Electric potential is a scalar quantity, that is, it has only magnitude and not direction. It may be viewed as analogous to temperature: as there is a certain temperature at every point in space, and the temperature gradient indicates the direction and magnitude of the driving force behind heat flow, similarly, there is an electric potential at every point in space, and its gradient, or field strength, indicates the direction and magnitude of the driving force behind charge movement. Equally, electric potential may be seen as analogous to height: just as a released object will fall through a difference in heights caused by a gravitational field, so a charge will 'fall' across the voltage caused by an electric field.

The electric field was formally defined as the force exerted per unit charge, but the concept of potential allows for a more useful and equivalent definition: the electric field is the local gradient of the electric potential. Usually expressed in volts per metre, the vector direction of the field is the line of greatest gradient of potential.

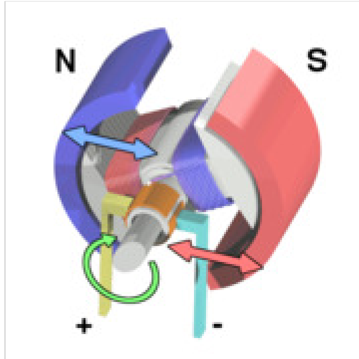
Electromagnetism



Magnetic field circles around a current

Ørsted's discovery in 1821 that a magnetic field existed around all sides of a wire carrying an electric current indicated that there was a direct relationship between electricity and magnetism. Moreover, the interaction seemed different to gravitational and electrostatic forces, the two forces of nature then known. The force on the compass needle did not direct it to or away from the current-carrying wire, but acted at right angles to it. Ørsted's slightly obscure words were that "the electric conflict acts in a revolving manner." The force also depended on the direction of the current, for if the flow was reversed, then the force did too.

Ørsted did not fully understand his discovery, but he observed the effect was reciprocal: a current exerts a force on a magnet, and a magnetic field exerts a force on a current. The phenomenon was further investigated by Ampère, who discovered that two parallel current carrying wires exerted a force upon each other: two wires conducting currents in the same direction are attracted to each other, while wires containing current flowing in opposite directions are forced apart. The interaction is mediated by the magnetic field each current produces and forms the basis for the international definition of the ampere.



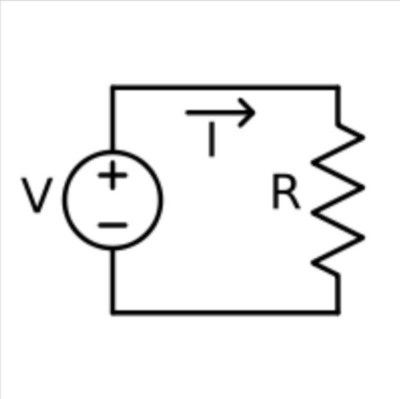
The electric motor exploits an important effect of electromagnetism: a current flowing through a magnetic field experiences a force at right angles to both the field and current

This relationship between magnetic fields and currents is extremely important, for it led to Michael Faraday's invention of the electric motor in 1821. Faraday's homopolar motor consisted of a permanent magnet sitting in a pool of mercury. A current was allowed to flow through a wire suspended from a pivot above the magnet and dipped into the mercury. The magnet exerted a tangential force on the wire, making it circle around the magnet for as long as the current was maintained.

Experimentation by Faraday in 1831 revealed that a wire moving perpendicular to a magnetic field developed a potential difference between its ends. Further analysis of this process, known as electromagnetic induction, enabled him to state the principal, now known as Faraday's law of induction, that the potential difference induced in a closed circuit is proportional to the rate of change of magnetic flux through the loop. Exploitation of this discovery enabled him to invent the first electrical generator in 1831, in which he converted the mechanical energy of a rotating copper disc to electrical energy. Faraday's disc was inefficient and of no use as a practical generator, but it showed the possibility of generating electric power using magnetism, a possibility that would be taken up by those that followed on from his work.

Faraday's and Ampère's work showed that a time-varying magnetic field acted as a source of an electric field, and a time-varying electric field was a source of a magnetic field. Thus, when either field is changing in time, then a field of the other is necessarily induced. Such a phenomenon has the properties of a wave, and is naturally referred to as an electromagnetic wave. Electromagnetic waves were analysed theoretically by James Clerk Maxwell in 1864. Maxwell discovered a set of equations that could unambiguously describe the interrelationship between electric field, magnetic field, electric charge, and electric current. He could moreover prove that such a wave would necessarily travel at the speed of light, and thus light itself was a form of electromagnetic radiation. Maxwell's Laws, which unify light, fields, and charge are one of the great milestones of theoretical physics.

Electric circuits



A basic electric circuit. The voltage source *V* on the left drives a current *I* around the circuit, delivering electrical energy into the resistance *R*. From the resistor, the current returns to the source, completing the circuit.

An electric circuit is an interconnection of electric components, usually to perform some useful task, with a return path to enable the charge to return to its source.

The components in an electric circuit can take many forms, which can include elements such as resistors, capacitors, switches, transformers and electronics. Electronic circuits contain active components, usually semiconductors, and typically exhibit non-linear behaviour, requiring complex analysis. The simplest electric components are those that are termed passive and linear: while they may temporarily store energy, they contain no sources of it, and exhibit linear responses to stimuli.

The resistor is perhaps the simplest of passive circuit elements: as its name suggests, it resists the flow of current through it, dissipating its energy as heat. Ohm's law is a basic law of circuit theory, stating that the current passing through a resistance is directly proportional to the potential difference across it. The ohm, the unit of resistance, was named in honour of Georg Ohm, and is symbolised by the Greek letter Ω. 1 Ω is the resistance that will produce a potential difference of one volt in response to a current of one amp.

The capacitor is a device capable of storing charge, and thereby storing electrical energy in the resulting field. Conceptually, it consists of two conducting plates separated by a thin insulating layer; in practice, thin metal foils are coiled together, increasing the surface area per unit volume and therefore the capacitance. The unit of capacitance is the farad, named after Faraday, and given the symbol *F*: one farad is the capacitance that develops a potential difference of one volt when it stores a charge of one coulomb. A capacitor connected to a voltage supply initially causes a current to flow as it accumulates charge; this current will however decay in time as the capacitor fills, eventually falling to zero. A capacitor will therefore not permit a steady-state current to flow, but instead blocks it.

The inductor is a conductor, usually a coil of wire, that stores energy in a magnetic field in response to the current flowing through it. When the current changes, the magnetic field does too, inducing a voltage between the ends of the conductor. The induced voltage is proportional to the time rate of change of the current. The constant of proportionality is termed the inductance. The unit of inductance is the henry, named after Joseph Henry, a contemporary of Faraday. One henry is the inductance that will induce a potential difference of one volt if the current through it changes at a rate of one ampere per second. The inductor's behaviour is in some regards converse to that of the capacitor: it will freely allow an unchanging current to flow, but opposes the flow of a rapidly changing one.

Production and uses

Generation

Wind power is of increasing importance in many countries

Thales' experiments with amber rods were the first studies into the production of electrical energy. While this method, now known as the triboelectric effect, is capable of lifting light objects and even generating sparks, it is extremely inefficient. It was not until the invention of the voltaic pile in the eighteenth century that a viable source of electricity became available. The voltaic pile, and its modern descendant, the electrical battery, store energy chemically and make it available on demand in the form of electrical energy. The battery is a versatile and very common power source which is ideally suited to many applications, but its energy storage is finite, and once discharged it must be disposed of or recharged. For large electrical demands electrical energy must be generated and transmitted in bulk.

Electrical energy is usually generated by electro-mechanical generators driven by steam produced from fossil fuel combustion, or the heat released from nuclear reactions; or from other sources such as kinetic energy extracted from wind or flowing water. Such generators bear no resemblance to Faraday's homopolar disc generator of 1831, but they still rely on his electromagnetic principle that a conductor linking a changing magnetic field induces a potential difference across its ends. The invention in the late nineteenth century of the transformer meant that electricity could be generated at centralised power stations, benefiting from economies of scale, and be transmitted across countries with increasing efficiency. Since electrical energy cannot easily be stored in quantities large enough to meet demands on a national scale, at all times exactly as much must be produced as is required. This requires electricity utilities to make careful predictions of their electrical loads, and maintain constant co-ordination with their power stations. A certain amount of generation must always be held in reserve to cushion an electrical grid against inevitable disturbances and losses.

Demand for electricity grows with great rapidity as a nation modernises and its economy develops. The United States showed a 12% increase in demand during each year of the first three decades of the twentieth century, a rate of growth that is now being experienced by emerging economies such as those of India or China. Historically, the growth rate for electricity demand has outstripped that for other forms of energy, such as coal.

Environmental concerns with electricity generation have led to an increased focus on generation from renewable sources, in particular from wind- and hydropower. While debate can be expected to continue over the environmental impact of different means of electricity production, its final form is relatively clean.

Uses

Electricity is an extremely flexible form of energy, and it may be adapted to a huge, and growing, number of uses. The invention of a practical incandescent light bulb in the 1870s led to lighting becoming one of the first publicly available applications of electrical power. Although electrification brought with it its own dangers, replacing the naked flames of gas lighting greatly reduced fire hazards within homes and factories. Public utilities were set up in many cities targeting the burgeoning market for electrical lighting.


The light bulb, an early application of electricity, operates by Joule heating: the passage of current through resistance generating heat

The Joule heating effect employed in the light bulb also sees more direct use in electric heating. While this is versatile and controllable, it can be seen as wasteful, since most electrical generation has already required the production of heat at a power station. A number of countries, such as Denmark, have issued legislation restricting or banning the use of electric heating in new buildings. Electricity is however a highly practical energy source for refrigeration, with air conditioning representing a growing sector for electricity demand, the effects of which electricity utilities are increasingly obliged to accommodate.

Electricity is used within telecommunications, and indeed the electrical telegraph, demonstrated commercially in 1837 by Cooke and Wheatstone, was one of its earliest applications. With the construction of first intercontinental, and then transatlantic, telegraph systems in the 1860s, electricity had enabled communications in minutes across the globe. Optical fibre and satellite communication technology have taken a share of the market for communications systems, but electricity can be expected to remain an essential part of the process.

The effects of electromagnetism are most visibly employed in the electric motor, which provides a clean and efficient means of motive power. A stationary motor such as a winch is easily provided with a supply of power, but a motor that moves with its application, such as an electric vehicle, is obliged to either carry along a power source such as a battery, or by collecting current from a sliding contact such as a pantograph, placing restrictions on its range or performance.

Electronic devices make use of the transistor, perhaps one of the most important inventions of the twentieth century, and a fundamental building block of all modern circuitry. A modern integrated circuit may contain several billion miniaturised transistors in a region only a few centimetres square.

Electricity and the natural world

Physiological effects

A voltage applied to a human body causes an electric current to flow through the tissues, and although the relationship is non-linear, the greater the voltage, the greater the current. The threshold for perception varies with the supply frequency and with the path of the current, but is about 1 mA for mains-frequency electricity. If the current is sufficiently high, it will cause muscle contraction, fibrillation of the heart, and tissue burns. The lack of any visible sign that a conductor is electrified makes electricity a particular hazard. The pain caused by an electric shock can be intense, leading electricity at times to be employed as a method of torture. Death caused by an electric shock is referred to as electrocution. Electrocution is still the means of judicial execution in some jurisdictions, though its use has become rarer in recent times.

Electrical phenomena in nature

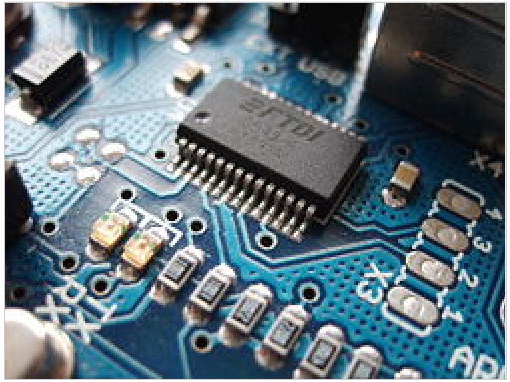

The electric eel, *Electrophorus electricus*

Electricity is by no means a purely human invention, and may be observed in several forms in nature, a prominent manifestation of which is lightning. The Earth's magnetic field is thought to arise from a natural dynamo of circulating currents in the planet's core. Certain crystals, such as quartz, or even cane sugar, generate a potential difference across their faces when subjected to external pressure. This phenomenon is known as piezoelectricity, from the Greek *piezein*, meaning to press, and was discovered in 1880 by Pierre and Jacques Curie. The effect is reciprocal, and when a piezoelectric material is subjected to an electric field, a small change in physical dimensions take place.

Some organisms, such as sharks, are able to detect and respond to changes in electric fields, an ability known as electroreception, while others, termed electrogenic, are able to generate voltages themselves to serve as a predatory or defensive weapon. The order Gymnotiformes, of which the best known example is the electric eel, detect or stun their prey via high voltages generated from modified muscle cells called electrocytes. All animals transmit information along their cell membranes with voltage pulses called action potentials, whose functions include communication by the nervous system between neurons and muscles. They are also responsible for coordinating activities in certain plants.

Electronics

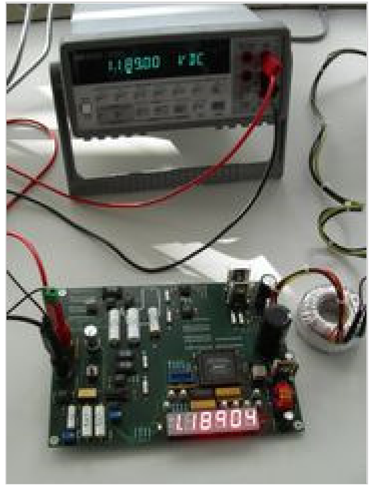
2008/9 Schools Wikipedia Selection. Related subjects: Engineering


Surface mount electronic components

Electronics is the study of the flow of charge through various materials and devices such as semiconductors, resistors, inductors, capacitors, nano-structures and vacuum tubes. Although considered to be a theoretical branch of physics, the design and construction of electronic circuits to solve practical problems is an essential technique in the fields of electronic engineering and computer engineering. This science starts about 1908 with the invention by Dr Lee De Forest of the valve (triode) Before 1950 this science was named "Radio" or "Radio technics" because that was its principal application.

The study of new semiconductor devices and surrounding technology is sometimes considered a branch of physics. This article focuses on engineering aspects of electronics.

Overview of electronic systems and circuits


Commercial digital voltmeter checking a prototype

Electronic systems are used to perform a wide variety of tasks. The main uses of electronic circuits are:

1. The controlling and processing of data.
2. The conversion to/from and distribution of electric power.

Both these applications involve the creation and/or detection of electromagnetic fields and electric currents. While electrical energy had been used for some time prior to the late 19th century to transmit data over telegraph and telephone lines, development in electronics grew exponentially after the advent of radio.

One way of looking at an electronic system is to divide it into 3 parts:

- Inputs – Electronic or mechanical sensors (or transducers). These devices take signals/information from external sources in the physical world (such as antennas or technology networks) and convert those signals/information into current/voltage or digital (high/low) signals within the system.

- Signal processors – These circuits serve to manipulate, interpret and transform inputted signals in order to make them useful for a desired application. Recently, complex signal processing has been accomplished with the use of Digital Signal Processors.

- Outputs – Actuators or other devices (such as transducers) that transform current/voltage signals back into useful physical form (e.g., by accomplishing a physical task such as rotating an electric motor).

For example, a television set contains these 3 parts. The television's input transforms a broadcast signal (received by an antenna or fed in through a cable) into a current/voltage signal that can be used by the device. Signal processing circuits inside the television extract information from this signal that dictates brightness, colour and sound level. Output devices then convert this information back into physical form. A cathode ray tube transforms electronic signals into a visible image on the screen. Magnet-driven speakers convert signals into audible sound.

Electronic devices and components

An electronic component is any physical entity in an electronic system whose intention is to affect the electrons or their associated fields in a desired manner consistent with the intended function of the electronic system. Components are generally intended to be in mutual electromechanical contact, usually by being soldered to a printed circuit board (PCB), to create an electronic circuit with a particular function (for example an amplifier, radio receiver, or oscillator). Components may be packaged singly or in more or less complex groups as integrated circuits.

Types of circuits

Analog circuits

Most analog electronic appliances, such as radio receivers, are constructed from combinations of a few types of basic circuits. Analog circuits use a continuous range of voltage as opposed to discrete levels as in digital circuits. The number of different analog circuits so far devised is huge, especially because a 'circuit' can be defined as anything from a single component, to systems containing thousands of components.

Hitachi J100 adjustable frequency drive chassis.

Analog circuits are sometimes called linear circuits although many non-linear effects are used in analog circuits such as mixers, modulators, etc. Good examples of analog circuits include vacuum tube and transistor amplifiers, operational amplifiers and oscillators.

Some analog circuitry these days may use digital or even microprocessor techniques to improve upon the basic performance of the circuit. This type of circuit is usually called "mixed signal."

Sometimes it may be difficult to differentiate between analog and digital circuits as they have elements of both linear and non-linear operation. An example is the comparator which takes in a continuous range of voltage but puts out only one of two levels as in a digital circuit. Similarly, an overdriven transistor amplifier can take on the characteristics of a controlled switch having essentially two levels of output.

Digital circuits

Digital circuits are electric circuits based on a number of discrete voltage levels. Digital circuits are the most common physical representation of Boolean algebra and are the basis of all digital computers. To most engineers, the terms "digital circuit", "digital system" and "logic" are interchangeable in the context of digital circuits. In most cases the number of different states of a node is two, represented by two voltage levels labeled "Low"(0) and "High"(1). Often "Low" will be near zero volts and "High" will be at a higher level depending on the supply voltage in use.

Computers, electronic clocks, and programmable logic controllers (used to control industrial processes) are constructed of digital circuits. Digital Signal Processors are another example.

Building-blocks:

- Logic gates
- Adders
- Binary Multipliers
- Flip-Flops
- Counters
- Registers
- Multiplexers
- Schmitt triggers

Highly integrated devices:

- Microprocessors
- Microcontrollers
- Application-specific integrated circuit(ASIC)
- Digital signal processor (DSP)
- Field-programmable gate array (FPGA)

Mixed-signal circuits

Mixed-signal circuits refers to integrated circuits (ICs) which have both analog circuits and digital circuits combined on a single semiconductor die or on the same circuit board. Mixed-signal circuits are becoming increasingly common. Mixed circuits are usually used to control an analog device using digital logic, for example the speed of a motor. Analog to digital converters and digital to analog converters are the primary examples. Other examples are transmission gates and buffers.

Heat dissipation and thermal management

Heat generated by electronic circuitry must be dissipated to prevent immediate failure and improve long term reliability. Techniques for heat dissipation can include heatsinks and fans for air cooling, and other forms of computer cooling such as water cooling. These techniques use convection, conduction, & radiation of heat energy.

Noise

Noise is associated with all electronic circuits. Noise is defined as unwanted disturbances superposed on a useful signal that tend to obscure its information content. Noise is not the same as signal distortion caused by a circuit.

Electronics theory

Mathematical methods are integral to the study of electronics. To become proficient in electronics it is also necessary to become proficient in the mathematics of circuit analysis.

Circuit analysis is the study of methods of solving generally linear systems for unknown variables such as the voltage at a certain node or the current though a certain branch of a network. A common analytical tool for this is the SPICE circuit simulator.

Also important to electronics is the study and understanding of electromagnetic field theory.

Electronic test equipment

Electronic test equipment is used to create stimulus signals and capture responses from electronic Devices Under Test (DUTs). In this way, the proper operation of the DUT can be proven or faults in the device can be traced and repaired.

Practical electronics engineering and assembly requires the use of many different kinds of electronic test equipment ranging from the very simple and inexpensive (such as a test light consisting of just a light bulb and a test lead) to extremely complex and sophisticated such as Automatic Test Equipment.

Computer aided design (CAD)

Today's electronics engineers have the ability to design circuits using premanufactured building blocks such as power supplies, semiconductors (such as transistors), and integrated circuits. Electronic design automation software programs include schematic capture programs and printed circuit board design programs. Popular names in the EDA software world are NI Multisim, Cadence (ORCAD), Eagle PCB and Schematic, Mentor (PADS PCB and LOGIC Schematic), Altium (Protel), LabCentre Electronics (Proteus) and many others.

Construction methods

Many different methods of connecting components have been used over the years. For instance, early electronics often used point to point wiring with components attached to wooden breadboards to construct circuits. Cordwood construction and wire wraps were other methods used. Most modern day electronics now use printed circuit boards (made of FR4), and highly integrated circuits. Health and environmental concerns associated with electronics assembly have gained increased attention in recent years, especially for products destined to the European Union, with its Restriction of Hazardous Substances Directive (RoHS) and Waste Electrical and Electronic Equipment Directive (WEEE), which went into force in July 2006.

Electronics industry

- Semiconductor sales leaders by year

Branch pages

- Digital electronics
- Analogue electronics
- Microelectronics
- Fuzzy electronics

- Circuit Design
- Integrated circuit
- Optoelectronics
- Semiconductor
- Semiconductor device

Engineering

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Engineering is the discipline and profession of applying scientific knowledge and utilizing natural laws and physical resources in order to design and implement materials, structures, machines, devices, systems, and processes that realize a desired objective and meet specified criteria. The American Engineers' Council for Professional Development (ECPD, the predecessor of ABET) has defined engineering as follows:

> "[T]he creative application of scientific principles to design or develop structures, machines, apparatus, or manufacturing processes, or works utilizing them singly or in combination; or to construct or operate the same with full cognizance of their design; or to forecast their behaviour under specific operating conditions; all as respects an intended function, economics of operation and safety to life and property."

One who practices engineering is called an **engineer**, and those licensed to do so may have more formal designations such as Professional Engineer, Chartered Engineer, or Incorporated Engineer. The broad discipline of engineering encompasses a range of more specialized subdisciplines, each with a more specific emphasis on certain fields of application and particular areas of technology.

History

The *concept* of engineering has existed since ancient times as humans devised fundamental inventions such as the pulley, lever, and wheel. Each of these inventions is consistent with the modern definition of engineering, exploiting basic mechanical principles to develop useful tools and objects.

The Watt steam engine, a major driver in the industrial revolution, underscores the importance of Engineering in modern history. This model is on display at the main building of the ETSIIM in Madrid, Spain

The term *engineering* itself has a much more recent etymology, deriving from the word *engineer*, which itself dates back to 1325, when an *engine'er* (literally, one who operates an *engine*) originally referred to "a constructor of military engines." In this context, now obsolete, an "engine" referred to a military machine, *i. e.*, a mechanical contraption used in war (for example, a catapult). The word "engine" itself is of even older origin, ultimately deriving from the Latin *ingenium* (c. 1250), meaning "innate quality, especially mental power, hence a clever invention."

Later, as the design of civilian structures such as bridges and buildings matured as a technical discipline, the term civil engineering entered the lexicon as a way to distinguish between those specializing in the construction of such non-military projects and those involved in the older discipline of military engineering (the original meaning of the word "engineering," now largely obsolete, with notable exceptions that have survived to the present day such as military engineering corps, *e. g.*, the U. S. Army Corps of Engineers).

The Acropolis and the Parthenon in Greece, the Roman aquaducts, Via Appia and the Colosseum, the Hanging Gardens of Babylon, the Pharos of Alexandria, the pyramids in Egypt, Teotihuacán and the cities and pyramids of the Mayan, Inca and Aztec Empires, the Great Wall of China, among many others, stand as a testament to the ingenuity and skill of the ancient civil and military engineers.

The earliest civil engineer known by name is Imhotep. As one of the officials of the Pharaoh, Djosèr, he probably designed and supervised the construction of the Pyramid of Djoser (the Step Pyramid) at Saqqara in Egypt around 2630- 2611 BC. He may also have been responsible for the first known use of columns in architecture.

The first electrical engineer is considered to be William Gilbert, with his 1600 publication of De Magnete, who was the originator of the term "electricity".

The first steam engine was built in 1698 by mechanical engineer Thomas Savery. The development of this device gave rise to the industrial revolution in the coming decades, allowing for the beginnings of mass production.

With the rise of engineering as a profession in the nineteenth century the term became more narrowly applied to fields in which mathematics and science were applied to these ends. Similarly, in addition to military and civil engineering the fields then known as the mechanic arts became incorporated into engineering.

Electrical Engineering can trace its origins in the experiments of Alessandro Volta in the 1800s, the experiments of Michael Faraday, Georg Ohm and others and the invention of the electric motor in 1872. The work of James Maxwell and Heinrich Hertz in the late 19th century gave rise to the field of Electronics. The later inventions of the vacuum tube and the transistor further accelerated the development of Electronics to such an extent that electrical and electronics engineers currently outnumber their colleagues of any other Engineering specialty.

The inventions of Thomas Savery and the Scottish engineer James Watt gave rise to modern Mechanical Engineering. The development of specialized machines and their maintenance tools during the industrial revolution led to the rapid growth of Mechanical Engineering both in its birthplace Britain and abroad.

Even though in its modern form Mechanical engineering originated in Britain, its origins trace back to early antiquity where ingenuous machines were developed both in the civilian and military domains. The Antikythera mechanism, the earliest known model of a mechanical computer in history, and the mechanical inventions of Archimedes, including his death ray, are examples of early mechanical engineering. Some of Archimedes' inventions as well as the Antikythera mechanism required sophisticated knowledge of differential gearing or epicyclic gearing, two key principles in machine theory that helped design the gear trains of the Industrial revolution and are still widely used today in diverse fields such as robotics and automotive engineering.

Chemical Engineering, like its counterpart Mechanical Engineering, developed in the nineteenth century during the Industrial Revolution. Industrial scale manufacturing demanded new materials and new processes and by 1880 the need for large scale production of chemicals was such that a new industry was created, dedicated to the development and large scale manufacturing of chemicals in new industrial plants. The role of the chemical engineer was the design of these chemical plants and processes.

Aeronautical Engineering deals with aircraft design while Aerospace Engineering is a more modern term that expands the reach envelope of the discipline by including spacecraft design. Its origins can be traced back to the aviation pioneers around the turn of the century from the 19th century to the 20th although the work of Sir George Cayley has recently been dated as being from the last decade of the 18th century. Early knowledge of aeronautical engineering was largely empirical with some concepts and skills imported from other branches of engineering. Only a decade after the successful flights by the Wright brothers, the 1920s saw extensive development of aeronautical engineering through development of World War I military aircraft. Meanwhile, research to provide fundamental background science continued by combining theoretical physics with experiments.

The first PhD in engineering (technically, *applied science and engineering*) awarded in the United States went to Willard Gibbs at Yale University in 1863; it was also the second PhD awarded in science in the U.S.

In 1990, with the rise of computer technology, the first search engine was built by computer engineer Alan Emtage.

Main Branches of Engineering

Engineering, much like science, is a broad discipline which is often broken down into several sub-disciplines. These disciplines concern themselves with differing areas of engineering work. Although initially an engineer will be trained in a specific discipline, throughout an engineer's career the engineer may become multi-disciplined, having worked in several of the outlined areas. Historically the main Branches of Engineering are categorized as follows:

- Aerospace Engineering - The design of aircraft, spacecraft and related topics.
- Chemical Engineering - The conversion of raw materials into usable commodities.
- Civil Engineering - The design and construction of public and private works, such as infrastructure, bridges and buildings.
- Electrical Engineering - The design of electrical systems, such as transformers, as well as electronic goods.
- Mechanical Engineering - The design of physical or mechanical systems, such as engines, powertrains, kinematic chains and vibration isolation equipment.

With the rapid advancement of Technology many new fields are gaining prominence and new branches are developing such as Computer Engineering, Software Engineering, Nanotechnology, Molecular engineering, Mechatronics etc. These new specialties sometimes combine with the traditional fields and form new branches such as Mechanical Engineering and Mechatronics and Electrical and Computer Engineering.

For each of these fields there exists considerable overlap, especially in the areas of the application of sciences to their disciplines such as physics, chemistry and mathematics.

Methodology

Design of a turbine requires collaboration from engineers from many fields

Engineers apply the sciences of physics and mathematics to find suitable solutions to problems or to make improvements to the status quo. If multiple options exist, engineers weigh different design choices on their merits and choose the solution that best matches the requirements. The crucial and unique task of the engineer is to identify, understand, and interpret the constraints on a design in order to produce a successful result. It is usually not enough to build a technically successful product; it must also meet further requirements. Constraints may include available resources, physical, imaginative or technical limitations, flexibility for future modifications and additions, and other factors, such as requirements for cost, safety, marketability, productibility, and serviceability. By understanding the constraints, engineers derive specifications for the limits within which a viable object or system may be produced and operated.

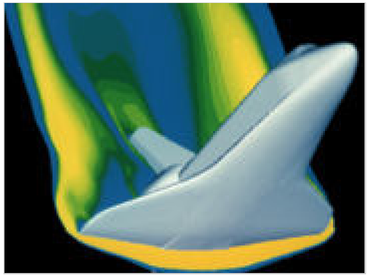
Problem solving

Engineers use their knowledge of science, mathematics, and appropriate experience to find suitable solutions to a problem. Engineering is considered a branch of applied mathematics and science. Creating an appropriate mathematical model of a problem allows them to analyze it (sometimes definitively), and to test potential solutions. Usually multiple reasonable solutions exist, so engineers must evaluate the different design choices on their merits and choose the solution that best meets their requirements. Genrich Altshuller, after gathering statistics on a large number of patents, suggested that compromises are at the heart of " low-level" engineering designs, while at a higher level the best design is one which eliminates the core contradiction causing the problem.

Engineers typically attempt to predict how well their designs will perform to their specifications prior to full-scale production. They use, among other things: prototypes, scale models, simulations, destructive tests, nondestructive tests, and stress tests. Testing ensures that products will perform as expected. Engineers as professionals take seriously their responsibility to produce designs that will perform as expected and will not cause unintended harm to the public at large. Engineers typically include a factor of safety in their designs to reduce the risk of unexpected failure. However, the greater the safety factor, the less efficient the design may be.

Computer use

As with all modern scientific and technological endeavors, computers and software play an increasingly important role. As well as the typical business application software there are a number of computer aided applications (CAx) specifically for engineering. Computers can be used to generate models of fundamental physical processes, which can be solved using numerical methods.


A computer simulation of high velocity air flow around the Space Shuttle during re-entry.

One of the most widely used tools in the profession is computer-aided design (CAD) software which enables engineers to create 3D models, 2D drawings, and schematics of their designs. CAD together with Digital mockup (DMU) and CAE software such as finite element method analysis allows engineers to create models of designs that can be analyzed without having to make expensive and time-consuming physical prototypes. These allow products and components to be checked for flaws; assess fit and assembly; study ergonomics; and to analyze static and dynamic characteristics of systems such as stresses, temperatures, electromagnetic emissions, electrical currents and voltages, digital logic levels, fluid flows, and kinematics. Access and distribution of all this information is generally organized with the use of Product Data Management software.

There are also many tools to support specific engineering tasks such as Computer-aided manufacture (CAM) software to generate CNC machining instructions; Manufacturing Process Management software for production engineering; EDA for printed circuit board (PCB) and circuit schematics for electronic engineers; MRO applications for maintenance management; and AEC software for civil engineering.

In recent years the use of computer software to aid the development of goods has collectively come to be known as Product Lifecycle Management (PLM).

Engineering in a social context

Engineering is a subject that ranges from large collaborations to small individual projects. Almost all engineering projects are beholden to some sort of financing agency: a company, a set of investors, or a government. The few types of engineering that are minimally constrained by such issues are pro bono engineering and open design engineering.

By its very nature engineering is bound up with society and human behaviour. Every product or construction used by modern society will have been influenced by engineering design. Engineering design is a very powerful tool to make changes to environment, society and economies, and its application brings with it a great responsibility, as represented by many of the Engineering Institutions codes of practice and ethics. Whereas medical ethics is a well-established field with considerable consensus, engineering ethics is far less developed, and engineering projects can be subject to considerable controversy. Just a few examples of this from different engineering disciplines are the development of nuclear weapons, the Three Gorges Dam, the design and use of Sports Utility Vehicles and the extraction of oil. There is a growing trend amongst western engineering companies to enact serious Corporate and Social Responsibility policies, but many companies do not have these.

Engineering is a key driver of human development. Sub-Saharan Africa in particular has a very small engineering capacity which results in many African nations being unable to develop crucial infrastructure without outside aid. The attainment of many of the Millennium Development Goals requires the achievement of sufficient engineering capacity to develop infrastructure and sustainable technological development. All overseas development and relief NGOs make considerable use of engineers to apply solutions in disaster and development scenarios. A number of charitable organizations aim to use engineering directly for the good of mankind:

- Engineers Without Borders
- Engineers Against Poverty
- Registered Engineers for Disaster Relief
- Engineers for a Sustainable World

Cultural presence

Engineering is a well respected profession. For example, in Canada it ranks as one of the public's most trusted professions.

Sometimes engineering has been seen as a somewhat dry, uninteresting field in popular culture, and has also been thought to be the domain of nerds. For example, the cartoon character Dilbert is an engineer. One difficulty in increasing public awareness of the profession is that average people, in the typical run of ordinary life, do not ever have any personal dealings with engineers, even though they benefit from their work every day. By contrast, it is common to visit a doctor at least once a year, the chartered accountant at tax time, and, occasionally, even a lawyer.

This has not always been so - most British school children in the 1950s were brought up with stirring tales of 'the Victorian Engineers', chief amongst whom were the Brunels, the Stephensons, Telford and their contemporaries.

In science fiction engineers are often portrayed as highly knowledgeable and respectable individuals who understand the overwhelming future technologies often portrayed in the genre. The *Star Trek* characters Montgomery Scott, Geordi La Forge, Miles O'Brien, B'Elanna Torres, and Charles Tucker are famous examples.

Occasionally, engineers may be recognized by the " Iron Ring"--a stainless steel or iron ring worn on the little finger of the dominant hand. This tradition began in 1925 in Canada for the Ritual of the Calling of an Engineer as a symbol of pride and obligation for the engineering profession. Some years later in 1972 this practice was adopted by several colleges in the United States. Members of the US Order of the Engineer accept this ring as a pledge to uphold the proud history of engineering. A Professional Engineer's name may be followed by the post-nominal letters PE or P.Eng in North America. In much of Europe a professional engineer is denoted by the letters IR, while in the UK and much of the Commonwealth the term Chartered Engineer applies and is denoted by the letters CEng.

Relationships with other disciplines

Science

> *Scientists study the world as it is; engineers create the world that has never been.*

— Theodore von Kármán

There exists an overlap between the sciences and engineering practice; in engineering, one applies science. Both areas of endeavor rely on accurate observation of materials and phenomena. Both use mathematics and classification criteria to analyze and communicate observations. Scientists are expected to interpret their observations and to make expert recommendations for practical action based on those interpretations. Scientists may also have to complete engineering tasks, such as designing experimental apparatus or building prototypes. Conversely, in the process of developing technology engineers sometimes find themselves exploring new phenomena, thus becoming, for the moment, scientists.

In the book *What Engineers Know and How They Know It*, Walter Vincenti asserts that engineering research has a character different from that of scientific research. First, it often deals with areas in which the basic physics and/or chemistry are well understood, but the problems themselves are too complex to solve in an exact manner. Examples are the use of numerical approximations to the Navier-Stokes equations to describe aerodynamic flow over an aircraft, or the use of Miner's rule to calculate fatigue damage. Second, engineering research employs many semi-empirical methods that are foreign to pure scientific research, one example being the method of parameter variation.

As stated by Fung et al. in the revision to the classic engineering text, Foundations of Solid Mechanics,

"Engineering is quite different from science. Scientists try to understand nature. Engineers try to make things that do not exist in nature. Engineers stress invention. To embody an invention the engineer must put his idea in concrete terms, and design something that people can use. That something can be a device, a gadget, a material, a method, a computing program, an innovative experiment, a new solution to a problem, or an improvement on what is existing. Since a design has to be concrete, it must have its geometry, dimensions, and characteristic numbers. Almost all engineers working on new designs find that they do not have all the needed information. Most often, they are limited by insufficient scientific knowledge. Thus they study mathematics, physics, chemistry, biology and mechanics. Often they have to add to the sciences relevant to their profession. Thus engineering sciences are born."

Medicine and biology

The study of the human body, albeit from different directions and for different purposes, is an important common link between medicine and some engineering disciplines. Medicine aims to sustain, enhance and even replace functions of the human body, if necessary, through the use of technology. Modern medicine can replace several of the body's functions through the use of artificial organs and can significantly alter the function of the human body through artificial devices such as, for example, brain implants and pacemakers. The fields of Bionics and medical Bionics are dedicated to the study of synthetic implants pertaining to natural systems. Conversely, some engineering disciplines view the human body as a biological machine worth studying, and are dedicated to emulating many of its functions by replacing biology with technology. This has led to fields such as artificial intelligence, neural networks, fuzzy logic, and robotics. There are also substantial interdisciplinary interactions between engineering and medicine.


Leonardo DaVinci, seen here in a self-portrait, has been described as the epitome of the artist/engineer. He is also known for his studies on human anatomy and physiognomy

Both fields provide solutions to real world problems. This often requires moving forward before phenomena are completely understood in a more rigorous scientific sense and therefore experimentation and empirical knowledge is an integral part of both. Medicine, in part, studies the function of the human body. The human body, as a biological machine, has many functions that can be modeled using Engineering methods. The heart for example functions much like a pump, the skeleton is like a linked structure with levers, the brain produces electrical signals etc. These similarities as well as the increasing importance and application of Engineering principles in Medicine, led to the development of the field of biomedical engineering that utilizes concepts developed in both disciplines.

Newly emerging branches of science, such as Systems biology, are adapting analytical tools traditionally used for engineering, such as systems modeling and computational analysis, to the description of biological systems.

Art

There are connections between engineering and art; they are direct in some fields, for example, architecture, landscape architecture and industrial design (even to the extent that these disciplines may sometimes be included in a University's Faculty of Engineering); and indirect in others. The Art Institute of Chicago, for instance, held an exhibition about the art of NASA's aerospace design. Robert Maillart's bridge design is perceived by some to have been deliberately artistic. At the University of South Florida, an engineering professor, through a grant with the National Science Foundation, has developed a course that connects art and engineering. Among famous historical figures Leonardo Da Vinci is a well known Renaissance artist and engineer, and a prime example of the nexus between art and engineering.

Other fields

In Political science the term *engineering* has been borrowed for the study of the subjects of Social engineering and Political engineering, which deal with forming political and social structures using engineering methodology coupled with political science principles.

Retrieved from " http://en.wikipedia.org/wiki/Engineering"

F-35 Lightning II

2008/9 Schools Wikipedia Selection. Related subjects: Air & Sea transport; Military History and War

F-35 Lightning II

The F-35 Lightning II takes off for its first flight at Naval Air Station Fort Worth Joint Reserve Base on 15 December 2006.

Type	Multirole fighter
Manufacturers	Lockheed Martin Aeronautics Northrop Grumman BAE Systems
Maiden flight	15 December 2006
Introduction	2011 (scheduled)
Status	Under development / pre-production
Primary users	United States Air Force United States Navy United States Marine Corps Royal Air Force/Royal Navy
Produced	2003-present

The **F-35 Lightning II** is a single-seat, single-engine, stealth-capable military strike fighter, a multi-role aircraft that can perform close air support, tactical bombing, and air-to-air combat. The F-35 is descended from the X-35 of the Joint Strike Fighter (JSF) program. Its development is being principally funded by the United States with the United Kingdom and other partner governments providing additional funding. It is being designed and built by an aerospace industry team led by Lockheed Martin with Northrop Grumman and BAE Systems as major partners. Demonstrator aircraft flew in 2000; a production model first took flight on 15 December 2006. The United States Air Force plans to acquire 1,763 aircraft.

Development

JSF Program history

Requirement

The JSF program was created to replace various aircraft while keeping development, production, and operating costs down. This was pursued by building three variants of one aircraft, sharing 80% of their parts:

Unit cost	US$200 million (flyaway cost based on producing 6 aircraft in 2008)
Developed from	Lockheed Martin X-35

- F-35A, conventional takeoff and landing (CTOL) variant.
- F-35B, short-takeoff and vertical-landing (STOVL) variant.
- F-35C, carrier-based (CV) variant.

The F-35 is being designed to be the world's premier strike aircraft through 2040. It is intended that its air-to-air capability will be second only to the F-22 Raptor. Specifically the F-35's requirements are that it be: four times more effective than legacy fighters in air-to-air combat, eight times more effective in air-to-ground battle combat, and three times more effective in reconnaissance and suppression of air defenses. These capabilities are to be achieved while still having significantly better range and a smaller logistical footprint than legacy aircraft.

An F-35 wind tunnel testing model in the Arnold Engineering Development Centre's 16-foot transonic wind tunnel

Origins and X-32 vs. X-35

The Joint Strike Fighter evolved out of several requirements for a common fighter to replace existing types. The actual JSF development contract was signed on 16 November 1996.

The contract for System Development and Demonstration (SDD) was awarded on 26 October 2001 to Lockheed Martin, whose X-35 beat the Boeing X-32. DoD officials and British Minister of Defence Procurement Lord Bach, said the X-35 consistently outperformed the X-32, although both met or exceeded requirements. The designation of the fighter as "F-35" came as a surprise to Lockheed, which had been referring to the aircraft in-house by the designation "F-24."

Boeing X-32 (left) and Lockheed Martin X-35 prior to down-select in 2001, where the X-35 was chosen.

Naming

On 7 July 2006, the U.S. Air Force officially announced the name of the F-35: Lightning II, in honour of Lockheed's World War II-era twin-prop P-38 Lightning and the Cold War-era jet, the English Electric Lightning. English Electric's aircraft division was incorporated into BAC, a predecessor of F-35 partner BAE Systems. Other names previously listed as contenders were Kestrel, Phoenix, Piasa, Black Mamba and Spitfire II. Lightning II was also an early company name for the aircraft that became the F-22 Raptor.

Design

The F-35 appears to be a smaller, slightly more conventional, one-engine sibling of the sleeker, two-engine F-22 Raptor, and indeed, drew elements from it. The exhaust duct design was inspired by the General Dynamics Model 200, a 1972 VTOL aircraft designed for the Sea Control Ship.

The F-35A being towed to its inauguration ceremony on 7 July 2006

Lockheed teamed with the Yakovlev Design Bureau in the 1990s, which has led to some speculation about ties with the quite different Yakovlev Yak-141 "Freestyle".

Stealth technology makes the aircraft hard to detect as it approaches short-range tracking radar.

Some improvements over current-generation fighter aircraft are:

- Durable, low-maintenance stealth technology;
- Integrated avionics and sensor fusion that combine information from off- and onboard sensors to increase the pilot's situational awareness and improve identification and weapon delivery, and to relay information quickly to other command and control (C2) nodes;
- High speed data networking including IEEE- 1394b and Fibre Channel.
- Low life-cycle costs.

Cockpit

The F-35 will feature a cockpit speech-recognition system (Direct Voice Input), improving the pilot's ability to operate the aircraft over the current-generation. The F-35 will be the first U.S. operational fixed-wing aircraft to use this system, although similar systems have been used in AV-8B and trialled in previous U.S. jets, particularly the F-16 VISTA. The system is integrated by Adacel Systems Inc with the speech recognition module supplied by SRI International

Although helmet-mounted displays have already been integrated into some fourth-generation fighters such as the Swedish JAS 39 Gripen, the F-35 will be the first in which helmet-mounted displays replace a head-up display altogether, also trialled in F-16 VISTA.

Sensors

The main sensor on board the F-35 is its AN/APG-81 AESA-radar, designed by Northrop Grumman Electronic Systems. It is augmented by the Electro-Optical Targeting System (EOTS) mounted under the nose of the aircraft, designed by Lockheed Martin and BAE. Further electro-optical sensors are distributed over the aircraft as part of the AN/AAS-37 system which acts as missile warning system and can aid in navigation and night operations.


EOTS under the nose of a mockup of the F-35.

Thrust-to-weight ratio

The F-35B variant was in danger of missing performance requirements because it weighed too much — reportedly, by 2,200 pounds (1,000 kg) or 8 percent. In response, Lockheed Martin added engine thrust and shed more than a ton by thinning the aircraft's skin; shrinking the weapons bay and vertical tails; rerouting some thrust from the roll-post outlets to the main nozzle; and redesigning the wing-mate joint, portions of the electrical system, and the portion of the aircraft immediately behind the cockpit. Weighing in at up to 60,000 lb the F-35 will be the heaviest aircraft of any type ever to fly with only one engine.

Manufacturing responsibilities

Lockheed Martin Aeronautics is the prime contractor and performs aircraft final assembly, overall system integration, mission system, and provides forward fuselage, wings and flight controls system. Northrop Grumman provides Active Electronically Scanned Array (AESA) radar, centre fuselage, weapons bay, and arrestor gear. BAE Systems provides Aft fuselage and empennages, horizontal and vertical tails, crew life support and escape systems, Electronic warfare systems, fuel system, and Flight Control Software (FCS1). Alenia will perform final assembly for Italy and, according to an Alenia executive, assembly of all European aircraft with the exception of the UK's.

Operational history

Testing

Image:Sdd f35test 009.jpg

The F-35A Lightning II's first flight on 15 December 2006

On 19 February 2006, the first F-35A (USAF version) was rolled out in Fort Worth, Texas. The aircraft underwent extensive ground testing at Naval Air Station Fort Worth Joint Reserve Base in fall 2006. On 15 September 2006 the first engine run of the F135 afterburning turbofan was conducted in an airframe, with the tests completed on 18 September after a static run with full afterburner. The engine runs were the first time that the F-35 was completely functional on its own power systems. On 15 December 2006, the F-35 completed its maiden flight.

On May 3, 2007, an electrical problem consisting of electrical arcing inside a hydraulic control box forced the aircraft to make an emergency landing. It was grounded until December 7th, when Test Pilot Jon Beesley flew a 55 minute test flight.

A unique feature of the test program is the use of the so-called Lockheed CATBird avionic testbed, a highly modified Boeing 737-330, inside of which are racks holding all of F-35's avionics, as well as a complete F-35 cockpit.

International participation

While the United States is the primary customer and financial backer, the United Kingdom, Italy, the Netherlands, Canada, Norway, Denmark, Australia and Turkey have contributed US$4.375 billion toward the development costs of the program. Total development costs are estimated at more than US$40 billion (underwritten largely by the United States), while the purchase of an estimated 2,400 planes is expected to cost an additional US$200 billion. The nine major partner nations plan to acquire over 3,100 F-35s through 2035, making the F-35 one of the most numerous jet fighters.

There are three levels of international participation. The levels generally reflect the financial stake in the program, the amount of technology transfer and subcontracts open for bid by national companies, and the order in which countries can obtain production aircraft. The United Kingdom is the sole "Level 1" partner, contributing US$2.5 billion, about 10% of the development costs under the 1995 Memorandum of Understanding that brought the UK into the project. Level 2 partners are Italy, which is contributing US$1 billion; and the Netherlands, US$800 million. Level 3 partners are Canada, US$440 million; Turkey, US$175 million; Australia, US$144 million; Norway, US$122 million; and Denmark, US$110 million. Israel and Singapore have joined as Security Cooperative Participants.

Some of the partner countries have wavered in their public commitment to the JSF program, hinting or warning that unless they receive more subcontracts or technology transfer, they will forsake JSF for the Eurofighter Typhoon, Saab Gripen, Dassault Rafale or simply upgrade their existing aircraft. Norway has several times threatened to put their support on hold unless substantial guarantees for an increased industrial share is provided. Despite this Norway has signed all the Memoranda of Understanding, including the latest one detailing the future production phase of the JSF program. They have, however, indicated that they will increase and strengthen their cooperation with both competitors of the JSF, the Typhoon and the Gripen.

United Kingdom

The United Kingdom plans to acquire 138 F-35s for its Royal Air Force and the Royal Navy.

The UK became increasingly frustrated by a lack of US commitment to grant access to the technology that would allow the UK to maintain and upgrade its F-35s without US involvement. This is understood to relate mainly to the software for the aircraft. For five years, British officials sought an ITAR waiver to secure greater technology transfer. This request, which has the blessing of the Bush administration, was repeatedly blocked by US Representative Henry Hyde, who says that the UK needs to tighten its laws protecting against the unauthorized transfer of the most advanced US technology to third parties.

BAE Systems CEO Mike Turner complained that the US had denied his company access to the aircraft's source code. On 21 December 2005, an article in the *Glasgow Herald* quoted the chairman of the House of Commons Defence Select Committee as saying "the UK might have to consider whether to continue in the programme" if no access were granted. Lord Drayson, Minister for Defence Procurement, took a firmer stance during a March 2006 visit to Washington: "We do expect the software technology transfer to take place. But if it does not take place we will not be able to purchase these aircraft," and he said there was a 'Plan B' if the deal fell through. This may have been the development of a navalized Typhoon.

On 27 May 2006, President George W. Bush and Prime Minister Tony Blair announced that "Both governments agree that the UK will have the ability to successfully operate, upgrade, employ, and maintain the Joint Strike Fighter such that the UK retains operational sovereignty over the aircraft." Despite this, concerns were still expressed about the lack of technology transfer as late as December 2006. Nevertheless, on 12 December 2006, Lord Drayson signed an agreement which met the UK's demands for further participation, i.e., access to software source codes and operational sovereignty. The agreement allows "an unbroken British chain of command" for operation of the aircraft. Drayson said Britain would "not be required to have a US citizen in our own operational chain of command". Drayson also said, however, that Britain is still considering an unspecified "Plan B" alternative to buying the Joint Strike Fighter.

On 25 July 2007, the Ministry of Defence confirmed that they have placed orders for the two new aircraft carriers of the Queen Elizabeth Class, that will allow the purchase of the F-35B variant.

Italy

Italy plans to acquire 109 F-35As for the Italian Air Force, and 22 F-35Bs for the Marina Militare (Italian Navy) to be used on the STOVL aircraft carrier Cavour.

Netherlands

The Netherlands has plans to acquire 85 F-35As for the Royal Netherlands Air Force. The aircraft will replace an aging fleet of Lockheed Martin F-16AM. The Dutch government expects the costs to be €5.5 billion for the initial purchase and €9.1 billion for 30 years of service. On 19 November 2007, in the Dutch Parliament, the Secretary of Defence was questioned about the JSF delay, technical problems and rising costs.

Australia

In May 2005, the Australian government announced that it would delay its planned 2006 decision on buying the JSF to 2008, and thus past the term of the government of the day. Australia, like the UK, has insisted it must have access to all software needed to modify and repair aircraft. Analysis conducted by the Royal Australian Air Force has determined that the F-35 "is the most suitable aircraft for Australia's needs".


Australia's then-Minister for Defence Dr. Brendan Nelson signing the JSF Production, Sustainment and Follow-on Development Memorandum of Understanding in December 2006

There has been debate in Australia over whether the F-35 is the most suitable aircraft for the RAAF. Some media reports, lobby groups and politicians have raised doubts that the aircraft will be ready in time to replace the RAAF's aging fleet of General Dynamics F-111 strike aircraft and F/A-18 Hornet fighters. Some critics say the more expensive F-22 or the Eurofighter may be better choices, both offering better range, dogfighting capability, and supercruise at a cost that may not be much more than the F-35 -- claims that as of July 2006 are being examined in a parliamentary inquiry.

In a statement released in early August 2006, then-Australian Defence Minister Dr. Brendan Nelson revealed that while the F-35 still had governmental support, Australia is starting to investigate other possible aircraft should the F-35 prove to be unfeasible. In October 2006 the deputy chief of the Air Force, Air Vice Marshal John Blackburn, publicly stated that the RAAF had ruled out the purchase of interim strike aircraft to cover any delays to the F-35 program and believed that the F-35 was suitable. However, on 6 March 2007, Dr. Nelson announced the Australian Government would purchase 24 F/A-18F Super Hornets from Boeing to fill the gap left by the retiring F-111 strike aircraft at a cost of potentially AU$6 billion. Nonetheless, Dr. Nelson continued to endorse Australia's purchase of the F-35. Speaking on Australian television in March 2007, Dr. Nelson stated that 5% of the capability of the F-35 is classified, claiming that, "that's the five percent that really counts."

On 13 December 2006, Minister Nelson signed the JSF Production, Sustainment and Follow-on Development Memorandum of Understanding. This agreement provides the cooperative framework for the acquisition and support of the JSF over its life. Australia is expected to purchase 100 F-35As at a cost of approximately AU$16 billion.

Turkey

On 12 July 2002, Turkey became the seventh international partner in the JSF Project, joining the United Kingdom, Italy, the Netherlands, Canada, Denmark and Norway. On 25 January 2007, Turkey signed a memorandum of understanding (MoU) for involvement in F-35 production. Turkey is expected to order 100 F-35A "CTOL/Air Force versions" at a reported cost of $11 billion. It is reported that the aircraft will be produced under license in Turkey by Turkish Aerospace Industries (TAI).

Turkish Production of the F-35

A Letter of Intent (LOI) was signed between TAI and Northrop Grumman ISS (NGISS) International on 6 February 2007. With the LOI, TAI becomes the second source for the F-35 Lightning II center fuselage during the JSF Signing. The number of centre fuselages to be produced by TAI will be determined depending on the number of F-35s Turkey will procure and the number of F-35s to be produced worldwide. The LOI represents a potential value in excess of $3 billion.

TAI of Turkey is one of the only two International Suppliers to Northrop Grumman Corporation (the other being Denmark). On 10 December, 2007 the Turkish Aerospace Industries, Inc. (TAI) was authorized by Northrop Grumman Corporation to commence fabricating subassemblies for the first two F-35 production aircraft. The subassemblies – composite components and aircraft access doors – will be used in the F-35 centre fuselage, a major section of the aircraft being produced by Northrop Grumman, a principal member of the Lockheed Martin-led F-35 global industry team.

In February 2007, Northrop Grumman also signed a letter of intent with TAI to also make it a second source for producing F-35 centre fuselages. Under that agreement, TAI will produce a *minimum* of 400 center fuselages starting during LRIP-2. Northrop Grumman currently produces all F-35 centre fuselages at its F-35 assembly facility in Palmdale, Calif.

It is also anticipated that TAI after 2013 will also produce 100% of the F-35 under licence from Lockheed Martin Corporation, as was also the case with the F-16 Fighting Falcon program Peace Onyx I and II.

Turkey also intends to incorporate in the distant future several Turkish designed and manufactured electronic systems into the F-35 platform.

Canada

The Canadian Department of National Defence (DND) is looking to replace its aging fleet of CF-18 Hornets by the 2020 time frame with much interest placed on the F-35. DND officials have stated the estimate for producing 80 units would cost $3.8 billion, though this figure does not include training, sustainment, and any follow on costs. Canada has until 2012 to decide on purchasing the F-35, though they have already invested $150 million in the JSF program.

Israel

In 2003, Israel signed a letter of agreement, worth almost $20 million, to formally join the system development and demonstration (SDD) effort for the F-35 as a "security cooperation participant" (SCP). The Israeli Air Force (IAF) stated in 2006 that the F-35 is a key part of IAF's recapitalization plans, and that Israel intends to buy over 100 F-35A fighters at an estimated cost of over $5 billion to replace their F-16s over time. Israel was reinstated as a partner in the development of the F-35 on 31 July 2006, after Israeli participation was put on hold following the Chinese arms deal crisis.

On 3 September 2007 IDF Chief of General Staff Lt.-Gen. Gabi Ashkenazi announced the purchase a squadron of F-35s which Israel will begin receiving in 2014. However, U.S. defense officials later agreed to allow Israel to receive the fighters as early as 2012. The price of each F-35 is expected to reach $50 million-$60m.

India

The F-35 is also a potential offer to the Indian Air Force as of July 2007. This has been interpreted as part of a tactic to sell the F-16 as a multi-role fighters to the IAF. India is not an official participant in the F-35 program.

Variants

The F-35 is planned to be built in three different versions to suit the needs of its various users.

F-35A

The F-35A, the conventional takeoff and landing (CTOL) variant intended for the US Air Force and other air forces. It is the smallest, lightest F-35 version and is the only variant equipped with an internal gun, the GAU-12/U. This 25 mm cannon, a development of the 20 mm M61 Vulcan carried by USAF fighters since the F-104 Starfighter, is also carried by the USMC's AV-8B Harrier II.

The F-35A not only matches the F-16 in maneuverability, instantaneous and sustained high-g performance, but also outperforms it in stealth, payload, range on internal fuel, avionics, operational effectiveness, supportability and survivability. It also has an internal laser designator and infra-red sensors.

It is primarily intended to replace the USAF's F-16 Fighting Falcons, beginning in 2013, and replace the A-10 Thunderbolt II starting in 2028.

F-35B

The F-35B is the short takeoff and vertical landing (STOVL) variant aircraft. The F-35B is similar in size to the Air Force F-35A, trading fuel volume for vertical flight systems. Like the AV-8 Harrier II, guns will be carried in a ventral pod. Vertical flight is by far the riskiest, and in the end, a decisive factor in design.

X-35B lift fan; the VTOL propulsion system is designed and manufactured by Rolls-Royce plc

The F-35's main power plant is derived from Pratt & Whitney's F119 or GE Rolls Royce fighter team's F136, with the STOVL variant of the latter incorporating a Rolls-Royce Lift Fan module. Instead of lift engines, or rotating nozzles on the engine fan and exhaust like the Pegasus-powered Harrier, the F-35B uses a vectoring cruise nozzle in the tail, i.e. the rear exhaust turns to deflect thrust down, and an innovative shaft-driven Lift Fan, patented by Lockheed Martin and developed by Rolls-Royce. Somewhat like a turboprop built within the fuselage, engine shaft power is diverted forward via a clutch-and- bevel gearbox to a vertically mounted, contra-rotating lift fan located forward of the main engine in the centre of the aircraft. Bypass air from the cruise engine turbofan exhausts through a pair of roll-post nozzles in the wings on either side of the fuselage, while the lift fan balances the vectoring cruise nozzle at the tail. This system is more similar to the Russian Yak-141 and German VJ 101D/E than previous STOVL designs, such as the Harrier with thrust vectoring.

In effect, the F-35B power plant acts as a flow multiplier, much as a turbofan achieves efficiencies by moving unburned air at a lower velocity, and getting the same effect as the Harrier's huge, but supersonically impractical, main fan. Like lift engines, this added machinery is dead weight during flight, but increased lifting power increases takeoff payload by even more. The cool fan exhaust also reduces the harmful effects of hot, high-velocity air which can harm runway pavement or an aircraft carrier deck. Though potentially risky and complicated, it was made to work to the satisfaction of DOD officials.

This variant is intended to replace the later derivatives of the Harrier Jump Jet, which was the world's first operational short takeoff, vertical landing fighter, ground attack aircraft. The RAF and Royal Navy will use this variant to replace the Harrier GR7/GR9s. The F-35B variant was unveiled at Lockheed's Fort Worth plant on December 18, 2007. The U.S. Marine Corps will use the F-35B to replace both its AV-8B Harrier II and F/A-18 Hornet fighters. The B variant is expected to be available beginning in 2012.

F-35C

The F-35C carrier variant will have a larger, folding wing and larger control surfaces for improved low-speed control, and stronger landing gear for the stresses of carrier landings. The larger wing area provides decreased landing speed, increased range and payload, with twice the range on internal fuel compared with the F/A-18C Hornet, achieving much the same goal as the heavier F/A-18E/F Super Hornet.

The US Navy intends to buy 480 F-35Cs to replace the F/A-18A, -B, -C, and -D Hornets. It will also serve as a stealthier complement to the Super Hornet. On 27 June 2007, the carrier variant completed its Air System Critical Design Review (CDR). This allows the F-35C to go to Low Rate Initial Production.

The C variant is expected to be available beginning in 2012.

Specifications (F-35 Lightning II)

Note: Some information is estimated.


The first of 15 pre-production F-35s

Data from F-35 Program brief, F-35 JSF Statistics

General characteristics

- **Crew:** 1
- **Length:** 50 ft 6 in (15.37 m)
- **Wingspan:** 35 ft 0 in (10.65 m)
- **Height:** 17 ft 4 in (5.28 m)
- **Wing area:** 459.6 ft² (42.7 m²)
- **Empty weight:** 26,000 lb (12,000 kg)
- **Loaded weight:** 44,400 lb (20,100 kg)
- **Max takeoff weight:** 60,000 lb (27,200 kg)
- **Powerplant:** 1× Pratt & Whitney F135 afterburning turbofan
  - **Dry thrust:** 25,000 lbf (111 kN)
  - **Thrust with afterburner:** >40,000 lbf (178 kN)
- **Secondary Powerplant**: 1× General Electric/Rolls-Royce F136 afterburning turbofan, >40,000 lbf (178 kN) [in development]
- **Lift fan** (STOVL): 1× Rolls-Royce Lift System driven from either F135 or F136 power plant, 18,000 lbf (80 kN)


A Pratt and Whitney F135 engine undergoes altitude testing at the Arnold Engineering Development Centre.

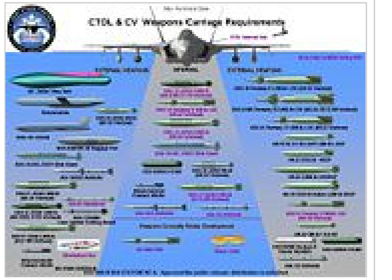
Performance

- **Maximum speed:** Mach 1.6+ (1,200 mph, 1,931 km/h)
- **Range:** A: 1,200 nmi; B: 900 nm; C: 1400 nm (A: 2,200 km; B: 1,667 km; C: 2,593 km) on internal fuel
- **Combat radius:** 600 nmi (690 mi, 1,110 km)
- **Rate of climb:** classified (not publicly available)
- **Wing loading:** 91.4 lb/ft² (446 kg/m²)
- **Thrust/weight:**
  - **With full fuel:** A: 0.89; B: 0.92; C: 0.81
  - **With 50% fuel:** A: 1.12; B: 1.10; C: 1.01

G-Limits

- F-35A: +9G
- F-35B: +7G
- F-35C: +7.5G

Armament


F-35A and F-35C armament

- 1 × GAU-12/U 25 mm cannon — slated to be mounted internally with 180 rounds in the F-35A and fitted as an external pod with 220 rounds in the F-35B and F-35C.

- Internally (current planned weapons for integration), up to six AIM-120 AMRAAM or AIM-132 ASRAAM or two air-to-air and two air-to-ground weapons (up to two 2,000 lb weapons in A and C models; two 1,000 lb weapons in the B model) in the bomb bays. These could be AMRAAM, the Joint Direct Attack Munition (JDAM) — up to 2,000 lb (910 kg), the Joint Standoff Weapon (JSOW), Small Diameter Bombs (SDB) — a maximum of four in each bay, the Brimstone anti-armor missiles, Cluster Munitions (WCMD) and High Speed Anti-Radiation Missiles (HARM). The MBDA Meteor air-to-air missile is currently being adapted to fit internally in the missile spots and may be integrated into the F-35.


Weapons bay on a mock-up of the F-35.

- At the expense of being more detectable by radar, many more missiles, bombs and fuel tanks can be attached on four wing pylons and two wingtip positions. The two wingtip pylons can only carry AIM-9 Sidewinders, while the Storm Shadow and Joint Air to Surface Stand-off Missile (JASSM) cruise missiles can be carried in addition to the stores already integrated. An air-to-air load of 12 AIM-120s and two AIM-9s is conceivable using internal and external weapons stations, as well as a configuration of six two thousand pound bombs, four AIM-120s and two AIM-9s.

Directed-energy weapons

Directed-energy weapons could be installed in some conventional takeoff F-35 Lightning IIs, whose lack of a direct lift fan frees up about 100 ft^3 (2.8 m^3) of space with access to a drive shaft capable of delivering more than 27,000 hp (20 MW). Some concepts, including solid-state lasers and high-power microwave beams, may be nearing operational status.

Popular culture

The first major film appearance of an F-35B was in *Live Free or Die Hard* (released as *Die Hard 4.0* or *Die Hard 4* outside North America) in 2007. The film used a combination of a full-scale model and CGI to dramatize its hovering ability using the lift fan.

Ford Motor Company

2008/9 Schools Wikipedia Selection. Related subjects: Road transport

Ford	
Type	Public (NYSE: F)
Founded	June 17, 1903
Founder	Henry Ford
Headquarters	Dearborn, Michigan, USA
Area served	worldwide
Key people	William Clay Ford, Jr - Executive Chairman Alan Mulally - President, CEO
Industry	Automotive
Products	Automobiles, Automotive parts
Services	Automotive financing and services
Revenue	▲ US$173.9 billion (2007)
Operating income	▲ US$126 million (2007)
Net income	▼ US$(-2.665) billion (2007)
Employees	▼ 245,000 (2007)
Divisions	Ford Credit Ford division Lincoln Mercury Premier Automotive Group

Ford Motor Company is an American multinational corporation and the world's third largest automaker based on worldwide vehicle sales.

In 2007, Ford became the third-ranked automaker in US sales after General Motors and Toyota, falling from the second-ranked automaker slot for the first time in the previous 56 years. Ford was also the overall seventh-ranked American-based company in the 2007 Fortune 500 list, based on global revenues in 2006 of $160.1 billion. In 2007, Ford revenues increased to $173.9 billion, while producing 6.553 million automobiles and employing about 245,000 employees at around 100 plants and facilities worldwide. Also in 2007, Ford received more quality survey awards from J. D. Power and Associates than any other automaker, with five vehicles ranking at the top of their categories, and fourteen vehicles ranked in the top three.

Based in Dearborn, Michigan, a suburb of Detroit, the automaker was founded by Henry Ford and incorporated in June 16, 1903. Ford now encompasses many global brands, including Lincoln and Mercury of the US, Jaguar and Land Rover of the UK, and Volvo of Sweden. Ford also owns a one-third controlling interest in Mazda.

Ford introduced methods for large-scale manufacturing of cars and large-scale management of an industrial workforce, especially elaborately engineered manufacturing sequences typified by moving assembly lines. Henry Ford's combination of highly efficient factories, highly paid workers, and low prices revolutionized manufacturing and came to be known around the world as Fordism by 1914.

History

Ford was launched in a converted factory in 1903 with $28,000 in cash from twelve investors, most notably John Francis Dodge and Horace Elgin Dodge who would later found the Dodge Brothers Motor Vehicle Company. During its early years, the company produced just a few cars a day at its factory on Mack Avenue in Detroit, Michigan. Groups of two or three men worked on each car from components made to order by other companies. Henry Ford was 40 years old when he founded the Ford Motor Company, which would go on to become one of the largest and most profitable companies in the world, as well as being one of the few to survive the Great Depression. The largest family-controlled company in the world, the Ford Motor Company has been in continuous family control for over 100 years.

Subsidiaries	Automotive Components Holdings Jaguar Land Rover Volvo (cars only)
Website	www.ford.com

Henry Ford (ca. 1919)

Corporate governance

Members of the board as of early 2007 are: Chief Sir John Bond, Richard Manoogian, Stephen Butler, Ellen Marram, Kimberly Casiano, Alan Mulally (President and CEO), Edsel Ford II, Homer Neal, William Clay Ford, Jr., Jorma Ollila, Irvine Hockaday, Jr., John L. Thornton and William Clay Ford (Director Emeritus).

The main corporate officers are: Lewis Booth (Executive Vice President, Chairman (PAG) and Ford of Europe), Mark Fields (Executive Vice President, President [The Americas]), Donat Leclair (Executive Vice President and CFO), Mark A. Schulz (Executive Vice President, President [International Operations]) and Michael E. Bannister (Group Vice President; Chairman & CEO Ford Motor Credit).. Paul Mascarenas (Vice President of Engineering, The Americas Product Development)

Recent company developments

1896 Ford Quadricycle

During the mid to late 1990s, Ford sold large numbers of vehicles, in a booming American economy with soaring stock market and low fuel prices. With the dawn of the new century, legacy healthcare costs, higher fuel prices, and a faltering economy led to falling market shares, declining sales, and sliding profit margins. Most of the corporate profits came from financing consumer automobile loans through Ford Motor Credit Company.

By 2005, corporate bond rating agencies had downgraded the bonds of both Ford and GM to junk status , citing high U.S. health care costs for an aging workforce, soaring gasoline prices, eroding market share, and dependence on declining SUV sales for revenues. Profit margins decreased on large vehicles due to increased "incentives" (in the form of rebates or low interest financing) to offset declining demand.

In the face of falling truck and SUV sales, Ford moved to introduce a range of new vehicles, including " Crossover SUVs" built on unibody car platforms, rather than body-on-frame truck chasses. Ford also developed alternative fuel and high efficiency vehicles, such as the Escape Hybrid.. Ford announced that it will team up with Southern California Edison (SCE) to examine the future of plug-in hybrids in terms of how home and vehicle energy systems will work with the electrical grid. Under the multi-million-dollar, multi-year project, Ford will convert a demonstration fleet of Ford Escape Hybrids into plug-in hybrids, and SCE will evaluate how the vehicles might interact with the home and the utility's electrical grid. Some of the vehicles will be evaluated "in typical customer settings," according to Ford.

In December 2006, the company raised its borrowing capacity to about $25 billion, placing substantially all corporate assets as collateral to secure the line of credit . Chairman Bill Ford has stated that "bankruptcy is not an option" , but economists have stated that the company's impending contract renewal with the United Auto Workers in the summer of 2007 could be brutal. The UAW has vowed to attempt to retain the jobs banks, a system which retains idled workers on the payroll, rather than laying them off, in order to maintain contracted US employment levels.

The automaker reported the largest annual loss in company history in 2006 of $12.7 billion, and estimated that it would not return to profitability until 2009. However, Ford surprised Wall Street in the second quarter of 2007 by posting a $750 million profit. The company finished the year with a $2.7 billion loss, largely attributed to finance restructuring at Volvo.

Ford has announced plans to sell Land Rover and Jaguar, with Tata Motors being the leading prospective buyer.

"The Way Forward"

In the latter half of 2005, Chairman Bill Ford asked newly-appointed Ford Americas Division President Mark Fields to develop a plan to return the company to profitability. Fields previewed the Plan, dubbed *The Way Forward*, at the December 7, 2005 board meeting of the company; and it was unveiled to the public on January 23, 2006. " The Way Forward" includes resizing the company to match current market realities, dropping some unprofitable and inefficient models, consolidating production lines, and shutting fourteen factories and cutting 30,000 jobs. .

These cutbacks are consistent with Ford's roughly 25% decline in U.S. automotive market share since the mid-late 1990s. Ford's target is to become profitable again in 2009, a year later than projected. Ford's realignment also includes the sale of its wholly owned subsidiary, Hertz Rent-a-Car to a private equity group for $15 billion in cash and debt acquisition. The sale was completed on December 22, 2005. A joint venture with Mahindra and Mahindra Limited of India ended with the sale of Ford's 15 percent stake in 2005.

Chairman and Chief Executive Officer Ford also became President of the company in April 2006, with the retirement of Jim Padilla. Five months later, in September, he stepped down as President and CEO, and naming Alan Mulally as his successor. Bill Ford continues as Executive Chairman, along with an executive operating committee made up of Mulally, Mark Schulz, Lewis Booth, Don Leclair, and Mark Fields.

Brands and marques

Today, Ford Motor Company manufactures automobiles under several names including Lincoln and Mercury in the United States. In 1958, Ford introduced a new marque, the Edsel, but poor sales led to its discontinuation in 1960. Later, in 1985, the Merkur brand was introduced; it met a similar fate in 1989.

Ford has major manufacturing operations in Canada, Mexico, the United Kingdom, Germany, Turkey, Brazil, Argentina, Australia, the People's Republic of China, and several other countries, including South Africa where, following divestment during apartheid, it once again has a wholly owned subsidiary. Ford also has a cooperative agreement with Russian automaker GAZ.

Since 1989, Ford has acquired Aston Martin (which it sold again on March 12, 2007, but it will retain a $77 million stake in the sports car maker), Jaguar, Land Rover, from the United Kingdom and Volvo Cars from Sweden, as well as a controlling share (33.4%) of Mazda of Japan, with which it operates an American joint venture plant in Flat Rock, Michigan called Auto Alliance. It has spun off its parts division under the name Visteon. Its prestige brands, with the exception of Lincoln, are managed through its Premier Automotive Group.

Ford's *FoMoCo* parts division sells aftermarket parts under the Motorcraft brand name.

Ford's non-manufacturing operations include organizations such as automotive finance operation Ford Motor Credit Company. Ford also sponsors numerous events and sports facilities around the nation, most notably Ford Centre in downtown Oklahoma City and Ford Field in downtown Detroit.

Overall the Ford Motor Company controls the following operational car marques: Daimler, Ford, Jaguar, Land Rover, Lincoln, Mazda, Mercury, and Volvo; Daimler, Jaguar, Land Rover, and Volvo are currently part of the Premier Automotive Group.

Global markets

Initially, Ford models sold outside the U.S. were essentially versions of those sold on the home market, but later on, models specific to Europe were developed and sold. Attempts to globalize the model line have often failed, with Europe's Ford Mondeo selling poorly in the United States, while U.S. models such as the Ford Taurus have fared poorly in Japan and Australia, even when produced in right hand drive. The small European model Ka, a hit in its home market, did not catch on in Japan, as it was not available as an automatic. The Mondeo was dropped by Ford Australia, because the segment of the market in which it competes had been in steady decline, with buyers preferring the larger local model, the Falcon. One recent exception is the European model of the Focus, which has sold strongly on both sides of the Atlantic.

From 2003, Toyota outsold Ford Motor worldwide. . From the second quarter 2006, Toyota has passed Ford as the #2 automaker, by sales, in the United States.

The Ford Motor Company is in partnership talks to license hybrid technology from the Toyota Motor Corporation in a deal that could help establish Toyota's system as a standard for the industry.

Europe

History

At first, Ford in Germany and the United Kingdom built different models from one another until the late 1960s, with the Ford Escort and then the Ford Capri being common to both companies. Later on, the Ford Taunus and Ford Cortina became identical, produced in left hand drive and right hand drive respectively. Rationalization of model ranges meant that production of many models in the UK switched to elsewhere in Europe, including Belgium and Spain as well as Germany. The Ford Sierra replaced the Taunus and Cortina in 1982, drawing criticism for its radical aerodynamic styling, which was soon given nicknames such as "Jellymould" and "The Salesman's Spaceship."

Increasingly, Ford Motor Company has looked to Ford of Europe for its "world cars," such as the Mondeo, Focus, and Fiesta, although sales of European-sourced Fords in the U.S. have been disappointing. In Asia, models from Europe are not as competitively priced as Japanese-built rivals, nor are they perceived as reliable. The Focus has been one exception to this, which has become America's best selling compact car since its launch in 2000.

In February 2002, Ford ended car production in the UK. It was the first time in 90 years that Ford cars had not been made in Britain, although production of the Transit van continues at the company's Southampton facility, engines at Bridgend and Dagenham, and transmissions at Halewood. Development of European Ford is broadly split between Dunton in Essex (powertrain, Fiesta/Ka, and commercial vehicles) and Cologne (body, chassis, electrical, Focus, Mondeo) in Germany. Ford also produced the Thames range of commercial vehicles, although the use of this brand name was discontinued circa 1965. It owns the Jaguar and/or Land Rover car plants in Britain; Ford's former Halewood Assembly Plant was converted for production of the Jaguar X-Type and currently also assembles Land-Rover's Freelander 2. Jaguars are also assembled at Castle Bromwich, Birmingham while the rest of the Land-Rover range is assembled at Solihull, near Birmingham.

Elsewhere in continental Europe, Ford assembles the Mondeo range in Genk (Belgium), Fiesta in Valencia (Spain) and Cologne (Germany), Ka in Valencia, and Focus in Valencia, Saarlouis (Germany) and Vsevolozhsk (Russia). Transit production is in Kocaeli (Turkey), Southampton (UK), and Transit Connect in Kocaeli.

Ford also owns a joint-venture production plant in Turkey. Ford-Otosan, established in the 1970s, manufactures the Transit Connect compact panel van as well as the "Jumbo" and long wheelbase versions of the full-size Transit. This new production facility was set up near Kocaeli in 2002, and its opening marked the end of Transit assembly in Genk.

Another joint venture plant near Setubal in Portugal, set up in collaboration with Volkswagen, formerly assembled the Galaxy people-carrier as well as its sister ships, the VW Sharan and Seat Alhambra. With the introduction of the third generation of the Galaxy, Ford has moved the production of the people-carrier to the Genk plant, with Volkswagen taking over sole ownership of the Setubal facility.

Ford Europe has broken new ground with a number of relatively futuristic car launches over the last 50 years.

Its 1959 Anglia two-door saloon was one of the most quirky-looking small family cars in Europe at the time of its launch, but buyers soon became accustomed to its looks and it was hugely popular with British buyers in particular. It was still selling well when replaced by the more practical Escort in 1967.

The third incarnation of the Ford Escort was launched in 1980 and marked the company's move from rear-wheel drive saloons to front-wheel drive hatchbacks in the small family car sector. It also offered levels of style, comfort and refinement which were almost unmatched on comparable cars of this era. It was a huge success all over Europe and it was Britain's most popular car for most of its 10-year production life.

The fourth generation Escort was produced from 1990 until 2000, although its successor - the Focus - had been on sale since 1998. On its launch, the Focus was arguably the most dramatic-looking and fine-handling small family cars on sale, and sold in huge volumes right up to the launch of the next generation Focus at the end of 2004.

The 1982 Ford Sierra - replacement for the long-running and massively popular Cortina and Taunus models - was a style-setter at the time of its launch. Its ultramodern aerodynamic design was a world away from a boxy, sharp-edged Cortina, and it was massively popular just about everywhere it was sold. A series of updates kept it looking relatively fresh until it was replaced by the front-wheel drive Mondeo at the start of 1993.

The first two incarnations of the Mondeo were well-built, refined and reliable family cars that attracted strong sales, but the third incarnation (launched in 2007) took the large family car market to new heights in terms of build quality, refinement, comfort, equipment, driver appeal and value for money.

The rise in popularity of small cars during the 1970s saw **Ford** enter the mini-car market in 1976 with its Fiesta hatchback. Most of its production was concentrated at Valencia in Spain, and the Fiesta sold in huge figures from the very start. An update in 1983 and the launch of an all-new model in 1989 strengthened its position in the small car market. The second generation Fiesta was significantly updated twice before an all-new model was launched in 2002, and over the years it has become more refined, spacious, better-built and more enjoyable to drive.

Asia Pacific

In New Zealand and Australia, the popular Ford Falcon was long considered the average family car and is considerably larger than the Mondeo, Ford's largest car sold in Europe. Between 1960 and 1972, the Falcon was based on a U.S. Ford of that name, but since then has been entirely designed and manufactured locally. Like its General Motors rival, the Holden Commodore, the 4.0 L Falcon retains rear wheel drive. High performance variants of the Falcon running locally-built engines produce up to 365 hp (272 kW). A ute (short for "utility," known in the US as pickup truck) version is also available with a similar range of drivetrains. In addition, Ford Australia sells highly-tuned Falcon sedans and utes through its performance car division, Ford Performance Vehicles. These cars produce 390 hp (291 kW) and are built in small numbers to increase their value as collectors' cars.

Ford dealership in Ho Chi Minh City, Vietnam (August 2005)

In Australia, the Commodore and Falcon have traditionally outsold all other cars and comprise over 20% of the new car market. In New Zealand, Ford was second in market share in the first eight months of 2006 with 14.4 per cent. This is all set to change with a shift away from local manufacturing and assembly: 2007 second quarter has seen Ford Australia cut their prestige (LWB) models and more recently, announced closure of their key engine manufacturing. This is due partly to drops in sales with stiff competition from Toyota's new Aurion and an updated Mitsubishi 380, both taking a large piece of the local family sedan market. Ford is betting on growth in small car sales with the Focus which it plans to assemble locally, and the popular Territory (Falcon-based) SUV.

Ford's presence in Asia has traditionally been much smaller. However, with the acquisition of a stake in Japanese manufacturer Mazda in 1979, Ford began selling Mazda's Familia and Capella (also known as the 323 and 626) as the Ford Laser and Telstar. The Laser was one of the most successful models sold by

Ford in Australia, and outsold the Mazda 323, despite being almost identical to it. The Laser was also built in Mexico and sold in the U.S. as the Mercury Tracer, while the 1991 (and on through the end of the model in the early 2000s) American Ford Escort (and 1991-on Tracer) was based on the Laser/Mazda 323, assembled in the US and Mexico.

Through its relationship with Mazda, Ford also acquired a stake in South Korean manufacturer Kia, which built the (Mazda-based) Ford Festiva from 1988-1993, and the Ford Aspire from 1994-1997 for export to the United States, but later sold their interest to Hyundai. Kia continued to market the Aspire as the Kia Avella, later replaced by the Rio and once again sold in the US. Ironically, Hyundai also manufactured the Ford Cortina until the 1980s. Ford also has a joint venture with Lio Ho in Taiwan, which assembled Ford models locally since the 1970s.

Ford came to India in 1998 with its Ford Escort model, which was later replaced by locally produced Ford Ikon in 2001. It has since added Fusion, Fiesta, Mondeo and Endeavour to its product line.

South America

In South America, Ford has had to face protectionist government measures in each country, with the result that it built different models in different countries, without particular regard to rationalization or economy of scale inherent to producing and sharing similar vehicles between the nations. In many cases, new vehicles in a country were based on those of the other manufacturers it had entered into production agreements with, or whose factories it had acquired. For example, the Corcel and Del Rey in Brazil were originally based on Renault vehicles.

In 1987, Ford merged its operations in Brazil and Argentina with those of Volkswagen to form a company called Autolatina, with which it shared models. Sales figures and profitability were disappointing, and Autolatina was dissolved in 1995. With the advent of Mercosur, the regional common market, Ford was finally able to rationalize its product line-ups in those countries. Consequently, the Ford Fiesta and Ford EcoSport are only built in Brazil, and the Ford Focus only built in Argentina, with each plant exporting in large volumes to the neighboring countries. Models like the Ford Mondeo from Europe could now be imported completely built up. Ford of Brazil produces a pick-up truck version of the Fiesta, the Courier, which is also produced in South Africa as the Ford Bantam in right hand drive versions.

Africa and Middle East

In Africa Ford's market presence has traditionally been strongest in South Africa and neighboring countries, with only trucks being sold elsewhere on the continent. Ford in South Africa began by importing kits from Canada to be assembled at its Port Elizabeth facility. Later Ford sourced its models from the UK and Australia, with local versions of the Ford Cortina including the XR6, with a 3.0 V6 engine, and a Cortina 'bakkie' or pick-up, which was exported to the UK. In the mid-1980s Ford merged with a rival company, owned by Anglo American, to form the South African Motor Corporation (Samcor).

Following international condemnation of apartheid, Ford divested from South Africa in 1988, and sold its stake in Samcor, although it licensed the use of its brand name to the company. Samcor began to assemble Mazdas as well, which affected its product line-up, which saw the European Fords like the Escort and Sierra replaced by the Mazda-based Laser and Telstar. Ford bought a 45 per cent stake in Samcor following the demise of apartheid in 1994, and this later became, once again, a wholly owned subsidiary, the Ford Motor Company of Southern Africa. Ford now sells a local sedan version of the Fiesta (also built in India and Mexico), and the Focus and Mondeo Europe. The Falcon model from Australia was also sold in South Africa, but was dropped in 2003.

Ford's market presence in the Middle East has traditionally been even smaller, partly due to previous Arab boycotts of companies dealing with Israel. Ford and Lincoln vehicles are currently marketed in ten countries in the region. Saudi Arabia, Kuwait, and the UAE are the biggest markets. Ford also established itself in Egypt in 1926, but faced an uphill battle during the 1950s due to the hostile nationalist business environment . Ford's distributor in Saudi Arabia announced in February 2003 that it had sold 100,000 Ford and Lincoln vehicles since commencing sales in November 1986. Half of the Ford/Lincoln vehicles sold in that country were Ford Crown Victorias. In 2004, Ford sold 30,000 units in the region, falling far short of General Motors' 88,852 units and Nissan Motors' 75,000 units.

Environmental record

Image:Bush-claycomo.jpg

Mulally (second from left) with President George W. Bush at the Kansas City Assembly plant in Claycomo, Missouri on March 20, 2007, touting Ford's new hybrid cars

Ford motor co. ranked 7th as one of the top corporate air polluters in the United States releasing 9.67 million pounds of toxic air and their toxic score was 244,782 in 2002. Some of the chemicals released were, Chromium 84 lb (38 kg), Formaldehyde 27042 lb (12266 kg), and Sulfuric Acid 5000 lb (2300 kg)

In 2000, under the leadership of the current Ford chairman, William Clay (Bill) Ford, the Company stunned the industry (and pleased environmentalists) with an announcement of a planned 25 percent improvement in the average mileage of its light truck fleet — including its popular SUVs — to be completed by the 2005 calendar year.

On the other hand, Ford ended the Think City experiment and ordered all the cars repossessed and destroyed, even as many of the people leasing them begged to be able to buy the cars from Ford. After outcry from the lessees and activists in the US and Norway, Ford returned the cars to Norway for sale.

In 2003, Ford announced that competitive market conditions and technological and cost challenges would prevent the company from achieving this goal. The US Environmental Protection Agency (EPA) released its 2005 fuel economy report ranking Ford cars, trucks and SUVs as having the lowest gas milleage of any automaker in America

Ford discontinued a line of electric Ranger pickup trucks and ordered them destroyed, though it reversed in January 2005, after environmentalist protest.

Ford did achieve significant progress toward improving fuel efficiency during 2005, with the successful introduction of the Hybrid-Electric Escape. The Escape's platform mate Mercury Mariner is also available with the hybrid-electric system in the 2006 model year—a full year ahead of schedule—due to high demand. The similar Mazda Tribute will also receive a hybrid-electric powertrain option, along with many other vehicles in the Ford vehicle line. In 2005, Ford announced its goal to make 250,000 hybrids a year by 2010, but by mid-2006 announced that it would not meet that goal. Other hybrids to come out will be the Ford Fusion and Mercury Milan Hybrid version in 2008. There are also plans for a Ford Edge and Lincoln MKX Hybrid. The Edge and MKX are Ford's new crossover SUVs to come out for the 2007 model year.

Ford also continues to study Fuel Cell-powered electric powertrains, and is currently demonstrating hydrogen-fueled internal combustion engine technologies, as well as developing the next-generation hybrid-electric systems. To the extent it is successful in increasing the percentage of hybrid vehicles and/or fuel cell vehicles, there will be a significant decrease not only of air pollution emissions but also reduced sound levels, with notable favorable impacts upon respiratory health and decrease of noise health effects.

While the company's product line increasingly reflects its commitment to ecologically sustainable practices, Ford's record as a manufacturer continues to reveal problematic ones. Researchers at the University of Massachusetts have listed it as the seventh-worst corporate producer of air pollution, primarily because of the manganese compounds, 1,2,4-trimethylbenzene, and glycol ethers released from its casting, truck, and assembly plants. The United States Environmental Protection Agency has linked Ford to 54 Superfund toxic waste sites, 12 of which have been cleaned up and deleted from the list.

Alternate fuel vehicles

Bill Ford was one of the first top industry executives to make regular use of an battery electric vehicle, a Ford Ranger EV, while the company contracted with the United States Postal Service to deliver electric postal vans based on the Ranger EV platform. The alternative fuel vehicles, such as some versions of the Crown Victoria especially in fleet and taxi service, operate on compressed natural gas - or CNG. Some CNG vehicles have dual fuel tanks - one for gasoline, the other for CNG - the same engine can operate on either fuel via a selector switch. Flexible fuel vehicles are designed to operate smoothly using a wide range of available fuel mixtures - from pure gasoline, to bioethanol-gasoline blends such as E85 (85% ethanol, 15% gasoline). Part of the challenge of successful marketing alternative and flexible fuel vehicles, is the general lack of establishment of sufficient fueling stations, which would be essential for these vehicles to be attractive to a wide range of consumers. Significant efforts to ramp up production and distribution of E85 fuels are underway and expanding.


2006 Ford Escape Hybrid

Current Ford Flexible Fuel Vehicles:

- Ford F-150
- Ford Crown Victoria
- Ford Focus / Focus C-MAX / Ford Focus FFV (Flexible-fuel vehicle).
- Ford Taurus
- Ford Ranger
- Ford Explorer
- Mercury Grand Marquis
- Lincoln Town Car


Ford Research Centre Aachen

Ford was third to the automotive market with a hybrid electric vehicle: the Ford Escape Hybrid, which also represented the first hybrid electric SUV] to market and started the Ford hybrid technology . The Hybrid Escape will also be the first hybrid electric vehicle with a Flexible Fuel capability to run on E85. The company had made plans to manufacture up to 250,000 hybrids a year by 2010, but has since had to back down on that commitment, due to excessively high costs and the lack of sufficient supplies of the hybrid-electric batteries and drivetrain system components. Instead, Ford has committed to accelerating development of next-generation hybrid-electric power plants in Britain, in collaboration with Volvo, Jaguar, and Land Rover. This engineering study is expected to yield more than 100 new hybrid-electric vehicle models and derivatives.

Ford is also planning to produce 250,000 E85-capable vehicles a year in the US, adding to some 1.6 million already sold in the last 10 years.

Ford also has launched the production of hydrogen-powered shuttle buses, using hydrogen instead of gasoline in a standard internal combustion engine, for use at airports and convention centers. At the 2006 Greater Los Angeles Auto Show, Ford showcased a hydrogen fuel cell version of its Explorer SUV. The Fuel cell Explorer has a combined output of 174 hp (130 kW). It has a large hydrogen storage tank which is situated in the centre of the car taking the original place of the conventional model’s automatic transmission. The centered position of the tank assists the vehicle reach a notable range of 350 miles (563 km), the farthest for a fuel cell vehicle so far. The fuel cell Explorer the first in a series of prototypes partly funded by the United States Department of Energy to expand efforts to determine the feasibility of hydrogen- powered vehicles. The fuel cell Explorer is one of several vehicles with green technology Ford being featured at the L.A. show, including the 2008 Ford Escape Hybrid, PZEV emissions compliant Fusion and Focus models and a 2008 Ford F-Series Super Duty outfitted with Ford's clean diesel technology.

Ford announced on 2007-07-09 that it will team up with Southern California Edison (SCE) to examine the future of plug-in hybrids in terms of how home and vehicle energy systems will work with the electrical grid. Under the multi-million-dollar, multi-year project, Ford will convert a demonstration fleet of Ford Escape Hybrids into plug-in hybrids, and SCE will evaluate how the vehicles might interact with the home and the utility's electrical grid. Some of the vehicles will be evaluated "in typical customer settings," according to Ford.

Current and planned Ford hybrid electric vehicles:

- 2004– Ford Escape Hybrid
- 2006– Mercury Mariner
- 2008– Ford Fusion/ Mercury Milan
- 2009– Ford Edge/ Lincoln MKX

Auto racing

NASCAR

NASCAR Ford Fusion race car

Ford is one of four manufacturers in the three NASCAR series: Nextel Cup, Busch Series, and Craftsman Truck Series. Major teams include Roush Fenway Racing and Yates Racing. Ford's racing teams debuted the Fusion race car, replacing the Taurus at the 2006 Daytona 500. Some of the most successful NASCAR Fords were the aerodynamic fastback Ford Torino and Mercury Montegos, and the aero-era Ford Thunderbirds.

Formula One


Rubens Barrichello driving for the Stewart Grand Prix team in 1997

Ford was heavily involved in Formula One for many years, and supplied engines to a large number of teams from 1967 until 2004. These engines were designed and manufactured by Cosworth, the racing division that was owned by Ford from 1998 to 2004. Ford-badged engines won 176 Grands Prix between 1967 and 2003 for teams such as Team Lotus and McLaren. Ford entered Formula One as a constructor in 2000 under the Jaguar Racing name, after buying the Stewart Grand Prix team which had been its primary 'works' team in the series since 1997. Jaguar achieved little success in Formula One, and after a turbulent five seasons, Ford withdrew from the category after the 2004 season, selling both Jaguar Racing (which became Red Bull Racing) and Cosworth (to Gerald Forsythe and Kevin Kalkhoven).

Rally

Marcus Grönholm driving the Ford Focus RS WRC 06 in 2006.

Ford has a long history in rallying and has been active in the World Rally Championship since the beginning of the world championship, the 1973 season. Ford took the 1979 manufacturers' title with Hannu Mikkola, Björn Waldegård and Ari Vatanen driving the Ford Escort RS1800. In the Group B era, Ford achieved success with Ford RS200. Since the 1999 season, Ford has used various versions of the Ford Focus WRC to much success. In the 2006 season, BP-Ford World Rally Team secured Ford its second manufacturers' title, with the Focus RS WRC 06 built by M-Sport and driven by *Flying Finns* Marcus Grönholm and Mikko Hirvonen. Continuing with Grönholm and Hirvonen, Ford successfully defended the manufacturers' world championship in the 2007 season. Ford is the only manufacturer to score in the points for 92 consecutive races; since the 2002 season opener Monte Carlo Rally.

Sports cars

Ford sports cars have always been visible in the world of endurance racing. Most notably the GT40 won the prestigious 24 Hours of Le Mans four times in the 1960s and still stands today as one of the all-time greatest racing cars. The GT40 is the only American car to ever win overall at Le Mans.

Ford won the manufacturers title in 2005 in the Grand-Am Cup series with the FR500C Mustang race car.

Touring cars

Ford has campaigned touring cars such as the Focus, Falcon, and Contour/ Mondeo and the Sierra Cosworth in many different series throughout the years. Notably, the Mondeo finished 1,2,3 in the British Touring Car Championship in 2000, and the Falcon finished 1,2,3 in the Australian V8 Supercar Series in 2005.

Other

Ford is the sole engine provider in the Champ Car series. The engines are manufactured by Cosworth. In the Indianapolis 500, Ford powered racing cars won 17 times between 1965 and 1996. Ford has a storied history in the Trans-Am series from the 1970s through today, having won many championships and races with its Ford Mustang. Ford has also branched out into drifting with the introduction of the new model mustang. Most noticeable is the Tourqoise and Blue Falken Tires Mustang driven by Vaughn Gittin Jr, A.K.A. "JR". with 750 RWHP (Rear Wheel Horsepower). In drag racing, John Force has piloted his Drag Ford Mustang to several NHRA funny-car titles in recent seasons. Formula Ford, a formula for single-seater cars without wings and originally on road tires were conceived in 1966 in the UK as an entry-level formula for racing drivers. Many of today's racing drivers started their car racing careers in this category.

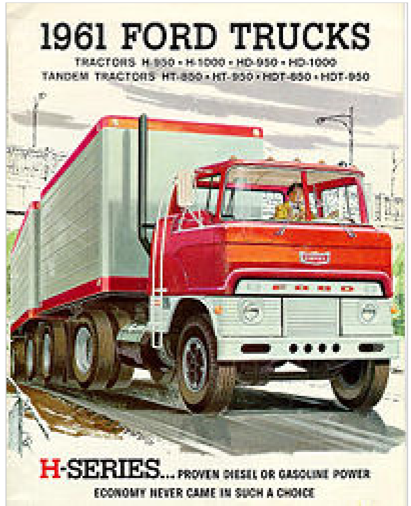
Ford trucks

1939 Ford pick-up truck

Ford has produced trucks since 1908. Countries where Ford commercial vehicles are or were made include Argentina, Australia, Brazil, Canada (badged Mercury too), France, Germany, India, Netherlands, Philippines, Spain (badged Ebro too), Turkey, UK (badged also Fordson and Thames) and USA.

Most of all these ventures are now extinct. The European one that lasted longer was the lorries arm of Ford of Britain, that was eventually sold to Iveco group in 1986, and whose last significant models were the Transcontinental and the Cargo.

In the USA, Ford's heavy trucks division (Classes 7 and 8) was sold in 1997 to Freightliner, now part of DaimlerChrysler, that rebranded it as Sterling. Ford continues building medium class trucks with the F-650 and F-750 and recently introduced the LCF series similar in design to the Ford Cargo trucks of the past.



1961 Ford H-Series trucks

Bus products

Ford has manufactured buses in the company's early history, but most Ford buses are built on Ford chassis by other manufacturers:

School Bus

- Ford Transit bus van chassis (formerly Econoline 350)
- Ford Minibus using F450
- Ford Minibus using E350
- Ford E450 Super Duty minibus

- Ford Class C School Bus using B600, B700, B800 chassis
- Ford MB IV 100, 100A, 200, 200C Super Duty

' Commercial Buses

- Ford MBC IV200
- Ford MBC IV 200C
- Ford MBC IV 300
- Ford MBC IV IV 300D
- Ford MBC II 800

Manufacturers using Ford E-series bodies include:

- Girardin Minibus Canada
- Corbeil Buses Canada

Commercial Bus

- Ford Specialty Trolley

Transit/Suburban Bus

- Ford G997
- Ford R1014
- Ford Trader
- Ford Hawke
- Ford ET7 with Casha bodywork
- Ford 19B, 29B
- Ford 72B
- Ford ET7 Aqualina

Clients include:

- Toronto Transportation Commission
- Kitchener Transit
- Hamilton Street Railway

Ford Tractors

The "Henry Ford and Son Company" began making Fordson tractors in Henry's hometown of Springwells (later part of Dearborn, Michigan from 1907 to 1928, from 1919 to 1932, at Cork, Ireland and 1933-1964 at Dagenham, England. They were also produced in Leningrad beginning in 1924.

In 1986, Ford expanded its tractor business when it purchased the Sperry-New Holland skid-steer loader and hay baler, hay tools and implement company from Sperry Corporation and formed Ford-New Holland which bought out Versatile tractors in 1988. This company was bought by Fiat and the name changed from Ford New Holland to New Holland. New Holland is now part of CNH Global.

Philco

Ford had non-automotive interests, as well, and owned the Philco home appliance company. Philco manufactured radios for Ford autos. See the Philco page for more information.

Criticism

Throughout its history, the company has faced a wide range of criticisms. Some have accused the early Fordist model of production of being exploitative, and Ford has been criticized as being willing to collaborate with dictatorships or hire mobs to intimidate union leaders and increase their profits through unethical means.

Ford refused to allow collective bargaining until 1941, with the Ford Service Department being set up as an internal security, intimidation, and espionage unit within the company, and quickly gained a reputation of using violence against union organizers and sympathizers.

Ford was also criticized for wearing down Firestone tires during driving, which caused many wrecks during a short time period in 2003. Many people were injured and killed due to the wearing down of the tires. Although Firestone received most of the blame, some blame fell on Ford, which advised customers to under-inflate the tires.

Alleged Nazi collaboration

Other accusations were that the company collaborated with the German Nazi regime and relied on Germany. The German Ford company used slave labor in Cologne between 1941 and 1945 and that it had produced military vehicles such as jeeps, planes, and ships used by a fascist regime. Many of these allegations were made in a series of United States lawsuits in 1998. The lawsuit was dismissed in 1999 because the judge concluded "the issues...concerned international treaties between nations and foreign policy and were thus in the realm of the executive branch."

Detractors point to Henry Ford's outspoken antisemitism, including his newspaper, The Dearborn Independent, which published The Protocols of the Elders of Zion, and "*The International Jew: The World's Foremost Problem*". Ford did not personally write "The International Jew", and later retracted it. They also point to the fact that in 1938, four months after the German annexation of Austria, Ford accepted the Grand Cross of the German Eagle, the Nazi regime's highest honour for foreigners before the outbreak of the war, as the only American ever to be given the award..

Defenders of the company argue that the Ford German division, Fordwerke, had been taken over by the Nazi government after it rose to power, claiming that it was not under the company's control, though Henry Ford, according to court records, did stay in touch with the company. Although Ford's initial motivations were anti-war, the company was heavily involved in the war effort after the outbreak of war.

Argentine "Dirty War"

Ford's Argentine subsidiary was accused of collaborating with the Argentine 1976-1983 military dictatorship, actively helping in the political repression of intellectuals and dissidents that was pursued by said government. No result was proven and the company denied the allegations.

In a lawsuit initiated in 1996 by relatives of some of the estimated 600 Spanish citizens who disappeared in Argentina during the " Dirty War", evidence was presented to support the allegation that much of this repression was directed by Ford and the other major industrial firms. According to a 5,000-page report, Ford executives drew up lists of "subversive" workers and handed them over to the military task-forces which were allowed to operate within the factories. These groups were allegedly kidnapped, tortured and murdered workers - at times allegedly within the plants themselves. The company denied the allegations.

In a second trial, a report brought by the CTA, and the testimonies of former Ford workers themselves, claimed that the company's Argentine factory was used between 1976 and 1978 as a detention centre, and that management allowed the military to set up its own bunker inside the plant. The company denied the allegations.

The Ford Pinto Memo

In September 1971 the Ford Motor Company launched the Pinto for the North American market. Through early production of this model it emerged that design flaws could result in fuel tank explosions when the vehicle was subject to a rear-end collision. Some sources even allege this safety data was available to Ford prior to production, but was ignored for economic reasons . Either way, a major scandal followed with the leaking to San Francisco magazine Mother Jones of the notorious "Ford Pinto Memo", an internal Ford cost-benefit analysis showing that the cost of implementing design changes to the subcompact's fuel system was greater than the economic cost of the burn injuries and deaths that could be prevented by doing so. Subsequently some have played down the importance of this case as Pinto explosion fatality estimates range widely from 27 to 900, with the lowest figures being allegedly in line with comparable fatality statistics for other car models. Nevertheless, the affair is an infamous example of a big corporation putting profit before human life because one senior Ford executive, at the time of the memo, is alleged to have written of his Pinto customers: it's "cheaper to let them burn" .

In the related Ford Pinto product liability case Grimshaw v. Ford Motor Co., 119 Cal. App. 3d 757 (4th Dist. 1981) the California Court of Appeal for the Fourth Appellate District reviewed Ford's conduct and upheld compensatory damages of $2.5 million and punitive damages of $3.5 million against Ford. Of the two plaintiffs, one was killed in the collision that caused her Pinto to explode, and her passenger, 13-year old Richard Grimshaw, was badly burned and scarred for life.

Glasses

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

"Glasses" can also be the plural of "glass."

Glasses, also called **eyeglasses** or **spectacles**, are frames bearing lenses worn in front of the eyes, normally for vision correction, eye protection, or for protection from UV rays.

A pair of modern glasses

Modern glasses are typically supported by pads on the bridge of the nose and by temples placed over the ears. Historical types include the pince-nez, monocle, and lorgnette.

Eyeglass frames are commonly made from metal or plastic. Lenses were originally made from glass, but many are now made from various types of plastic, including CR-39 or polycarbonate. These materials reduce the danger of breakage and weigh less than glass lenses. Some plastics also have more advantageous optical properties than glass, such as better transmission of visible light and greater absorption of ultraviolet light. Some plastics have a greater index of refraction than most types of glass; this is useful in the making of corrective lenses shaped to correct various vision abnormalities such as myopia, allowing thinner lenses for a given prescription.

French Empire gilt scissors glasses c.1805

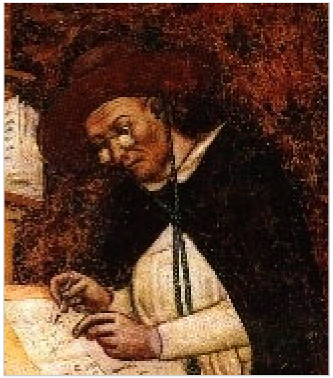
Scratch-resistant coatings can be applied to most plastic lenses giving them similar scratch resistance to glass. Hydrophobic coatings designed to ease cleaning are also available, as are anti-reflective coatings intended to improve night vision and make the wearer's eyes more visible.

Polycarbonate lenses are the lightest and most shatter-resistant, making them the best for impact protection, yet offer poor optics due to high dispersion, and having a low Abbe number of 31. CR-39 lenses are the most common plastic lenses due to their low weight, high scratch resistance, and low transparency for ultra violet and infrared radiation.

Not all glasses are designed solely for vision correction, but rather for protection, viewing visual information (such as stereoscopy) or simply just for aesthetic or fashion values. Safety glasses are a kind of eye protection against flying debris or against visible and near visible light or radiation. Sunglasses allow better vision in bright daylight, and may protect against damage from high levels of ultraviolet light.

History

Precursors



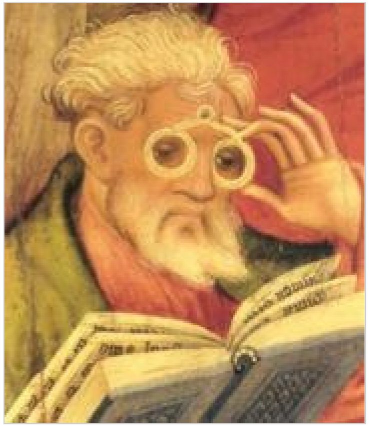
Detail of a portrait of Hugh de Provence, painted by Tomaso da Modena in 1352

The first suspected recorded use of a corrective lens may have been by the emperor Nero in the 1st century, who was known to watch the gladiatorial games using an emerald.

Corrective lenses were said to be used by Abbas Ibn Firnas in the 9th century. He had devised a way to finish sand into glass; which until this time, was secret to the Egyptians. These glasses could be shaped and polished into round rocks used for viewing - known as reading stones. Sunglasses, in the form of flat panes of smoky quartz, protected the eyes from glare and were used in China in the 12th century or possibly earlier. However, they did not offer any corrective powers.

According to "Great Hisatory" by *Beethoven*, some historians from USA found a blueprint of a glasses in Tokyo Japan, and the date on it was before the 4th century. The historians are right now examing the accurate date of the blueprint was first made. This means that the first person in the world who actually invent glasses may be a Japanese.

Invention of eyeglasses



The 'Glasses Apostle' by Conrad von Soest (1403)

Around 1284 in Italy, Salvino D'Armate is credited with inventing the first wearable eye glasses. The earliest pictorial evidence for the use of eyeglasses, however, is Tomaso da Modena's 1352 portrait of the cardinal Hugh de Provence reading in a scriptorium. Another early example would be a depiction of eyeglasses found north of the Alpes in an altarpiece of the church of Bad Wildungen, Germany, in 1403.

Many theories abound for to whom the credit for the invention of traditional eyeglasses belong. In 1676, Francesco Redi, a professor of medicine at the University of Pisa, wrote that he possessed a 1289 manuscript whose author complains that he would be unable to read or write were it not for the recent invention of glasses. He also produced a record of a sermon given in 1305, in which the speaker, a Dominican monk named Fra Giordano da Rivalto, remarked that glasses had been invented less than twenty years previously, and that he had met the inventor. Based on this evidence, Redi credited another Dominican monk, Fra Alessandro da Spina of Pisa, with the re-invention of glasses after their original inventor kept them a secret, a claim contained in da Spina's obituary record.

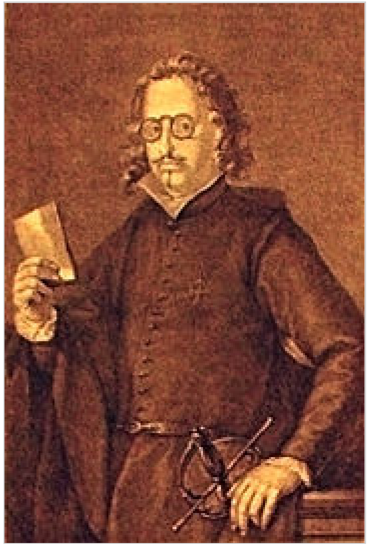
Other stories, possibly legendary, credit Roger Bacon with the invention. Bacon is known to have made the first recorded reference to the magnifying properties of lenses in 1262. His treatise *De iride* ("On the Rainbow"), which was written while he was a student of Robert Grosseteste, no later than 1235, mentions using optics to "read the smallest letters at incredible distances". While the exact date and inventor may be forever disputed, it is almost certainly clear that spectacles were invented between 1280 and 1300 in Italy.


Seated apostle holding lenses in position for reading. Detail from *Death of the Virgin*, by the Master of Heiligenkreuz, ca. 1400-30 (Getty Centre).

These early spectacles had convex lenses that could correct the presbyopia (farsightedness) that commonly develops as a symptom of aging. Nicholas of Cusa is believed to have discovered the benefits of concave lens in the treatment of myopia (nearsightedness). However, it was not until 1604 that Johannes Kepler published in his treatise on optics and astronomy, the first correct explanation as to why convex and concave lenses could correct presbyopia and myopia.

Later developments

The American scientist Benjamin Franklin, who suffered from both myopia and presbyopia, invented bifocals in 1784 to avoid having to regularly switch between two pairs of glasses. The first lenses for correcting astigmatism were constructed by the British astronomer George Airy in 1825.


A portrait of Francisco de Quevedo y Villegas, 1580–1645

Over time, the construction of spectacle frames also evolved. Early eyepieces were designed to be either held in place by hand or by exerting pressure on the nose (pince-nez). Girolamo Savonarola suggested that eyepieces could be held in place by a ribbon passed over the wearer's head, this in turn secured by the weight of a hat. The modern style of glasses, held by temples passing over the ears, was developed in 1727 by the British optician Edward Scarlett. These designs were not immediately successful, however, and various styles with attached handles such as " scissors-glasses" and lorgnettes remained fashionable throughout the 18th and into the early 19th century.

In the early 20th century, Moritz von Rohr at Zeiss (with the assistance of H. Boegehold and A. Sonnefeld), developed the Zeiss Punktal spherical point-focus lenses that dominated the eyeglass lens field for many years.

Despite the increasing popularity of contact lenses and laser corrective eye surgery, glasses remain very common and their technology has not stood still. For instance, it is now possible to purchase frames made of special memory metal alloys that return to their correct shape after being bent. Other frames have spring-loaded hinges. Either of these designs offers dramatically better ability to withstand the stresses of daily wear and the occasional accident. Modern frames are also often made from strong, light-weight materials such as titanium alloys, which were not available in earlier times.

On May 1, 1992 the United States Federal Trade Commission declared (section 456.2) that optometrists be required to provide the patient with a complete prescription immediately following an eye exam, effectively giving the patient the choice of where to purchase their glasses. The result was greater competition between the glasses manufacturers and thus lower prices for consumers. This trend has been accelerated by the proliferation of Internet technology, giving consumers the chance to bypass traditional distribution channels and buy glasses directly from the manufacturers.

Types

Corrective


Seattle skyline as seen through a corrective lens, showing the effect of refraction.

Corrective lenses modify the focal length of the eye to alleviate the effects of nearsightedness (myopia), farsightedness (hyperopia) or astigmatism. As people age, the eye's crystalline lens loses elasticity, resulting in presbyopia, which limits their ability to change focus.

The power of a lens is generally measured in diopters. Over-the-counter reading glasses are typically rated at +1.00 to +4.00 diopters. Glasses correcting for myopia will have negative diopter strengths. Lenses made to conform to the prescription of an ophthalmologist or optometrist are called prescription lenses and are used to make prescription glasses.

Safety

Safety glasses are usually made with shatter-resistant plastic lenses to protect the eye from flying debris. Although safety lenses may be constructed from a variety of materials of various impact resistance, certain standards suggest that they maintain a minimum 1 millimeter thickness at the thinnest point, regardless of material. Safety glasses can vary in the level of protection they provide. For example, those used in medicine may be expected to protect against blood splatter while safety glasses in a factory might have stronger lenses and a stronger frame with additional shields at the temples. The lenses of safety glasses can also be shaped for correction.


Safety glasses with side shields

The American National Standards Institute has established standard ANSI Z87.1 for safety glasses in the United States, and similar standards have been established elsewhere.

Some safety glasses are designed to fit over corrective glasses or sunglasses. They may provide less eye protection than goggles or other forms of eye protection, but their light weight increases the likelihood that they will actually be used. Recent safety glasses have tended to be given a more stylish design, in order to encourage their use. The pictured *wraparound safety glasses* are evidence of this style change with the close fitting nature of the wraparound dispensing with the need for side shields. Corrective glasses with plastic lenses can be used in the place of safety glasses in many environments; this is one advantage that they have over contact lenses.

There are also safety glasses for welding, which are styled like wraparound sunglasses, but with much darker lenses, for use in welding where a full sized welding helmet is inconvenient or uncomfortable. These are often called "flash goggles", because they provide protection from welding flash.

Nylon frames are usually used for protection eyewear for sports because of their lightweight and flexible properties. They are able to bend slightly and return to their original shape instead of breaking when pressure is applied to them. Nylon frames can become very brittle with age and they can be difficult to adjust.

Sunglasses

Scratch-resistant sunglasses made using a NASA developed coating

Sunglasses may be made with either prescription or non-prescription lenses that are darkened to provide protection against bright visible and possibly ultraviolet light.

Glasses with photosensitive lenses, called photochromic lenses, become darker in the presence of UV light. Unfortunately, many car windshields block the passage of UV light, making photochromic lenses less effective whilst driving on bright days. Still, they offer the convenience of not having to carry both clear glasses and sunglasses to those who frequently go indoors and outdoors during the course of a day.

Light polarization is an added feature that can be applied to sunglass lenses. Polarization filters remove horizontal rays of light, which can cause glare. Popular among fishermen and hunters, polarized sunglasses allow wearers to see into water when normally glare or reflected light would be seen. Polarized sunglasses may present some difficulties for pilots since reflections from water and other structures often used to gauge altitude may be removed, or instrument readings on liquid crystal displays may be blocked.

Yellow lenses are commonly used by golfers and shooters for their contrast enhancement and depth perception properties. Brown lenses are also common among golfers, but cause colour distortion. Blue, purple, and green lenses offer no real benefits to vision enhancement and are mainly cosmetic. Some sunglasses with interchangeable lenses have optional clear lenses to protect the eyes during low light or night time activities and a colored lens with UV protection for times where sun protection is needed. Debate exists as to whether "blue blocking" or amber tinted lenses have a protective effect.

Sunglasses are often worn just for aesthetic purposes, or simply to hide the eyes. Examples of sunglasses that were popular for these reasons include teashades and mirrorshades.

Special

The illusion of three dimensions on a two dimensional surface can be created by providing each eye with different visual information. Classic 3D glasses create the illusion of three dimensions when viewing specially prepared images. The classic 3D glasses have one red lens and one blue lens. 3D glasses made of cardboard and plastic are distributed at 3D movies. Another kind of 3D glasses uses polarized filters, with one lens polarized vertically and the other horizontally, with the two images required for stereo vision polarized the same way. The polarized 3D specs allow for colour 3D, while the red-blue lenses produce a dull black-and-white picture with red and blue fringes.


Swimming goggles.

One kind of electronic 3D spectacles uses electronic shutters, while virtual reality glasses and helmets have separate video screens for each eye.

Variations

Glasses can be very simple, such as magnifying lenses which are used to treat mild hyperopia and presbyopia can be bought off the shelf, normally referred to as reading glasses. Most glasses are made to a particular prescription, based on degree of myopia or hyperopia combined with astigmatism. Lenses can be ground to specific prescriptions, but in some cases standard off-the-shelf prescriptions suffice, but require custom fitting to particular frames.

As people age, their ability to focus is lessened and many decide to use multiple-focus lenses, bifocal or even trifocal to cover all the situations in which they use their sight. Traditional multifocal lenses have two or three distinct viewing areas, each requiring a conscious effort of refocusing. Some modern multifocal lenses, such as Progressive lenses (known as "no-line bifocals"), give a smooth transition between these different focal points and is unnoticeable by most wearers, while others have lenses specifically intended for use with computer monitors at a fixed distance. People may have several pairs of glasses, one for each task or distance, with specific glasses for reading, computer use, television watching, and writing.

Rimless

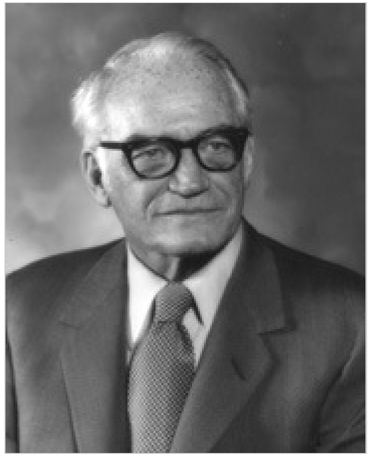
Three-piece rimless and semi-rimless glasses are common variations that differ from regular glasses in that their frames do not completely encircle the lenses. Three-piece rimless glasses have no frame around the lenses, and the bridge and temples are mounted directly onto the lenses. Semi-rimless (or half-rimless) glasses have a frame that only partially encircles the lenses (commonly the top portion), which are held in place most often by high strength nylon wire. A rare and currently non commercial variation are rimless and frameless glasses attached to a piercing at the bridge of a wearers nose. Such glasses have the visual look of the pince-nez.

Glazing

Spectacle lenses are edged into the frame's rim using glazing machines operated by ophthalmic technicians. The edging process begins with a trace being taken of the frame's eye shape. In earlier days the trace was replicated onto a plastic pattern called a Former. Nowadays the process is patternless and the shape is sent to the edger electronically.

The lens, in the form of a round uncut, is positioned in the correct manner to match the prescription and a block is stuck to the lens and that block fits into a chuck in the edging machine. A diamond coated wheel spins as the edger replicates the frame's eye-shape to the uncut lens. A 'v' bevel is applied to allow the edge of the lens to fit into the frame rim.

Fashion


United States senator Barry Goldwater in Horn-rimmed glasses.

Glasses can be a major part of personal expression, from the extravagance of Elton John and Dame Edna Everage, from Groucho Marx to Buddy Holly.

For some celebrities, glasses form part of their identity. American Senator Barry Goldwater continued to wear lensless horn-rimmed spectacles after being fitted with contact lenses because he was not recognizable without his trademark glasses. British soap star Anne Kirkbride had the same problem: her character on *Coronation Street*, Deirdre Barlow, became so well-known for her big frames that she was expected to wear them at social gatherings and in international tours, even though Kirkbride has always worn contact lenses. Drew Carey continued to wear glasses for the same reason after getting corrective laser eye surgery. British comedic actor Eric Sykes, who became profoundly deaf as an adult, wears glasses that contain no lenses; they are actually a bone-conducting hearing aid. Masaharu Morimoto wears glasses to separate his professional persona as a chef from his stage persona as Iron Chef Japanese. John Lennon wore his round-lens 'Windsor' spectacles from some of his time with the Beatles to his murder in 1980. Rock band Weezer are known for some of the members wearing thick-rimmed glasses.

In popular culture, glasses were all the disguise Superman and Wonder Woman needed to hide in plain view as alter egos Clark Kent and Diana Prince, respectively. An example of halo effect is seen in the stereotype that those who wear glasses are intelligent or, especially in teen culture, even geeks and nerds. Some people who find that wearing glasses may look nerdy turn to contact lenses instead, especially under peer pressure. Others turn to laser eye surgery, as do some would-be pilots.

Another unpopular aspect of glasses is their inconvenience. Even through the creation of light frames, such as those made of titanium, very flexible frames, and new lens materials and optical coatings, glasses can still cause problems during rigorous sports. The lenses can become greasy or trap vapour when eating hot food, swimming, walking in rain or rapid temperature changes (such as walking into a warm building from cold temperatures outside), reducing visibility significantly. Scraping, fracturing, or breakage of the lenses require time-consuming and costly professional repair, though modern plastic lenses are almost indestructible and very scratch-resistant.

Apple, Inc. co-founder Stephen Wozniak had a pair of eyeglasses made with lenses in the shape of the well-known Apple logo. The lenses were made from a block of acrylic, laminated from layers in the usual rainbow colors, and machined into the appropriate outline, with a custom-made frame in the same shape. They were made by a Silicon Valley optician.

Gold

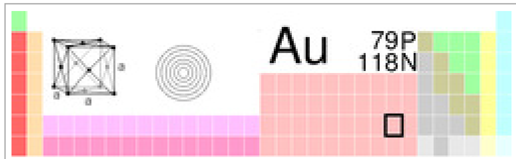
2008/9 Schools Wikipedia Selection. Related subjects: Chemical elements; Mineralogy

Gold (pronounced /ˈɡoʊld/) is a chemical element with the symbol **Au** (from the Latin *aurum*, meaning shining dawn) and atomic number 79. It is a highly sought-after precious metal which, for many centuries, has been used as money, a store of value and in jewelry. The metal occurs as nuggets or grains in rocks, underground "veins" and in alluvial deposits. It is one of the coinage metals. Gold is dense, soft, shiny and the most malleable and ductile of the known metals. Pure gold has a bright yellow colour traditionally considered attractive.

Gold formed the basis for the gold standard used before the fiat currency monetary system was employed by the International Monetary Fund (IMF) and the Bank for International Settlements (BIS). It is specifically against IMF regulations to base any currency against gold for all IMF member states. The ISO currency code of gold bullion is **XAU**.

Modern industrial uses include dentistry and electronics, where gold has traditionally found use because of its good resistance to oxidative corrosion.

Chemically, gold is a trivalent and univalent transition metal. Gold does not react with most chemicals, but is attacked by chlorine, fluorine, aqua regia and cyanide. Gold dissolves in mercury, forming amalgam alloys, but does not react with it. Gold is insoluble in nitric acid, which will dissolve silver and base metals, and this is the basis of the gold refining technique known as "inquartation and parting". Nitric acid has long been used to confirm the presence of gold in items, and this is the origin of the colloquial term "acid test," referring to a *gold standard* test for genuine value.

79	platinum ← gold → mercury
Ag ↑ **Au** ↓ Rg	 Periodic Table - Extended Periodic Table

General	
Name, Symbol, Number	gold, Au, 79
Chemical series	transition metals
Group, Period, Block	11, 6, d
Appearance	metallic yellow 
Standard atomic weight	196.966569 (4) $g \cdot mol^{-1}$
Electron configuration	[Xe] $4f^{14}$ $5d^{10}$ $6s^{1}$
Electrons per shell	2, 8, 18, 32, 18, 1
Physical properties	
Phase	solid

Characteristics

Gold is the most malleable and ductile metal; a single gram can be beaten into a sheet of one square meter, or an ounce into 300 square feet. Gold leaf can be beaten thin enough to become translucent. The transmitted light appears greenish blue, because gold strongly reflects yellow and red.

Gold readily forms alloys with many other metals. These alloys can be produced to increase the hardness or to create exotic colors (see below). Gold is a good conductor of heat and electricity, and is not affected by air and most reagents. Heat, moisture, oxygen, and most corrosive agents have very little chemical effect on gold, making it well-suited for use in coins and jewelry; conversely, halogens will chemically alter gold, and aqua regia dissolves it via formation of the chloraurate ion.

Common oxidation states of gold include +1 (gold(I) or aurous compounds) and +3 (gold(III) or auric compounds). Gold ions in solution are readily reduced and precipitated out as gold metal by adding any other metal as the reducing agent. The added metal is oxidized and dissolves allowing the gold to be displaced from solution and be recovered as a solid precipitate.

Recent research undertaken by Sir Frank Reith of the Australian National University shows that microbes play an important role in forming gold deposits, transporting and precipitating gold to form grains and nuggets that collect in alluvial deposits.

High quality pure metallic gold is tasteless, in keeping with its resistance to corrosion (it is metal ions which confer taste to metals).

In addition, gold is very dense, a cubic meter weighing 19300 kg. By comparison, the density of lead is 11340 kg/m³, and the densest element, iridium, is 22650 kg/m³.

Colour of gold

The usual gray colour of metals depends on their "electron sea" that is capable of absorbing and re-emitting photons over a wide range of frequencies. Gold behaves differently, depending on subtle relativistic effects that affect the orbitals around gold atoms.

Applications

As the metal

Medium of monetary exchange

In various countries, gold is used as a standard for monetary exchange, in coinage and in jewelry. Pure gold is too soft for ordinary use and is typically hardened by alloying with copper or other base metals. The gold content of gold alloys is measured in carats (k), pure gold being designated as 24k.

Gold coins intended for circulation from 1526 into the 1930s were typically a standard 22k alloy called

Density (near r.t.)	19.3 $g \cdot cm^{-3}$
Liquid density at m.p.	17.31 $g \cdot cm^{-3}$
Melting point	1337.33 K (1064.18 ° C, 1947.52 ° F)
Boiling point	3129 K (2856 ° C, 5173 ° F)
Heat of fusion	12.55 $kJ \cdot mol^{-1}$
Heat of vaporization	324 $kJ \cdot mol^{-1}$
Specific heat capacity	(25 °C) 25.418 $J \cdot mol^{-1} \cdot K^{-1}$

Vapor pressure

P(Pa)	1	10	100	1 k	10 k	100 k
at T(K)	1646	1814	2021	2281	2620	3078

Atomic properties

Crystal structure	cubic face centered
Oxidation states	−1, 1, 2, **3**, 4, 5 (amphoteric oxide)
Electronegativity	2.54 (Pauling scale)
Ionization energies	1st: 890.1 kJ/mol
	2nd: 1980 kJ/mol
Atomic radius	135 pm
Atomic radius (calc.)	174 pm
Covalent radius	144 pm
Van der Waals radius	166 pm

Miscellaneous

Magnetic ordering	diamagnetic
Electrical resistivity	(20 °C) 22.14 n Ω·m

crown gold, for hardness. Modern collector/investment bullion coins (which do not require good mechanical wear properties) are typically 24k, although the American Gold Eagle and British gold sovereign continue to be made at 22k, on historical tradition. The Canadian Gold Maple Leaf coin contains the highest purity gold of any popular bullion coin, at 99.999% (.99999 fine). Several other 99.99% pure gold coins are currently available, including Australia's Gold Kangaroos (first appearing in 1986 as the Australian Gold Nugget, with the kangaroo theme appearing in 1989), the several coins of the Australian Lunar Calendar series, and the Austrian Philharmonic. In 2006, the U.S. Mint began production of the American Buffalo gold bullion coin also at 99.99% purity.

Today, gold has fallen out of use in coins made for general circulation.

Jewelry

Because of the softness of pure (24k) gold, it is usually alloyed with base metals for use in jewelry, altering its hardness and ductility, melting point, colour and other properties. Alloys with lower "k", typically 22k, 18k, 14k or 10k, contain higher percentages of copper, silver or other base metals in the alloy. Copper is the most commonly used base metal, yielding a redder metal. Eighteen carat gold containing 25% copper is found in antique and Russian jewelry and has a distinct, though not dominant, copper cast, creating rose gold. Fourteen carat gold-copper alloy is nearly identical in colour to certain bronze alloys, and both may be used to produce police and other badges. Blue gold can be made by alloying with iron and purple gold can be made by alloying with aluminium, although rarely done except in specialized jewelry. Blue gold is more brittle and therefore more difficult to work with when making jewelry. Fourteen and eighteen carat gold alloys with silver alone appear greenish-yellow and are referred to as green gold. White gold alloys can be made with palladium or nickel. White 18 carat gold containing 17.3% nickel, 5.5% zinc and 2.2% copper is silver in appearance. Nickel is toxic, however, and its release from nickel white gold is controlled by legislation in Europe. Alternative white gold alloys are available based on palladium, silver and other white metals (World Gold Council), but the palladium alloys are more expensive than those using nickel. High-carat white gold alloys are far more resistant to corrosion than are either pure silver or sterling silver. The Japanese craft of Mokume-gane exploits the colour contrasts between laminated colored gold alloys to produce decorative wood-grain effects.

Thermal conductivity	(300 K) 318 $W{\cdot}m^{-1}{\cdot}K^{-1}$
Thermal expansion	(25 °C) 14.2 $\mu m{\cdot}m^{-1}{\cdot}K^{-1}$
Speed of sound (thin rod)	(r.t.) (hard-drawn) 2030 $m{\cdot}s^{-1}$
Young's modulus	78 GPa
Tensile Strain	0.00157
Shear modulus	27 GPa
Bulk modulus	220 GPa
Poisson ratio	0.44
Mohs hardness	2.5
Vickers hardness	216 MPa
Brinell hardness	? 2450 MPa
CAS registry number	7440-57-5

Selected isotopes

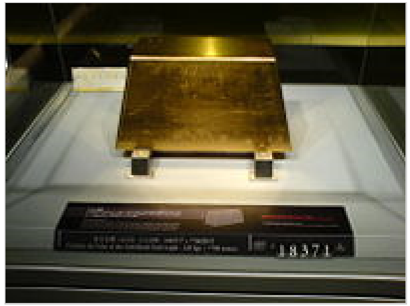
Main article: Isotopes of gold

iso	NA	half-life	DM	DE (MeV)	DP
^{195}Au	syn	186.10 d	ε	0.227	^{195}Pt
^{196}Au	syn	6.183 d	ε	1.506	^{196}Pt
			β^-	0.686	^{196}Hg
^{197}Au	100%	^{197}Au is stable with 118 neutrons			
^{198}Au	syn	2.69517 d	β^-	1.372	^{198}Hg
^{199}Au	syn	3.169 d	β^-	0.453	^{199}Hg

References

Other

- In medieval times, gold was often seen as beneficial for the health, in the belief that something that rare and beautiful could not be anything but healthy. Even some modern esotericists and forms of alternative medicine assign metallic gold a healing power. Some gold salts do have anti-inflammatory properties and are used as pharmaceuticals in the treatment of arthritis and other similar conditions. However, only salts and radioisotopes of gold are of pharmacological value, as elemental (metallic) gold is inert to all chemicals it encounters inside the body.
- Gold leaf, flake or dust is used on and in some gourmet foodstuffs, notably sweets and drinks as decorative ingredient. Gold flake was used by the nobility in Medieval Europe as a decoration in foodstuffs and drinks, in the form of leafs, flakes or dust, either to demonstrate the host's wealth or in the belief that something that valuable and rare must be beneficial for one's health.
- Gold solder is used for joining the components of gold jewelry by high-temperature hard soldering or brazing. If the work is to be of hallmarking quality, gold solder must match the carat weight of the work, and alloy formulas are manufactured in most industry-standard carat weights to colour match yellow and white gold. Gold solder is usually made in at least three melting-point ranges referred to as Easy, Medium and Hard. By using the hard, high-melting point solder first, followed by solders with progressively lower melting points, goldsmiths can assemble complex items with several separate soldered joints.
- Gold can be used in food and has the E Number 175. Goldwasser (German: "Goldwater") is a traditional herbal liqueur produced in Gdańsk, Poland and Schwabach, Germany and contains flakes of gold leaf. There are also some expensive (~$1000) cocktails which contain flakes of gold leaf. However, since metallic gold is inert to all body chemistry, it adds no taste nor has it any other nutritional effect and leaves the body unaltered.
- *Dentistry.* Gold alloys are used in restorative dentistry, especially in tooth restorations, such as crowns and permanent bridges. The gold alloys' slight malleability facilitates the creation of a superior molar mating surface with other teeth and produces results that are generally more satisfactory than those produced by the creation of porcelain crowns. The use of gold crowns in more prominent teeth such as incisors is favored in some cultures and discouraged in others.
- Gold can be made into thread and used in embroidery.
- Gold is ductile and malleable, meaning it can be drawn into very thin wire and can be beaten into very thin sheets known as gold leaf.
- Gold produces a deep, intense red colour when used as a coloring agent in cranberry glass.
- In photography, Gold toners are used to shift the colour of silver bromide black and white prints towards brown or blue tones, or to increase their stability. Used on sepia-toned prints, gold toners produce red tones. Kodak publish formulas for several types of gold toners, which use gold as the chloride (Kodak, 2006).
- *Electronics.* The concentration of free electrons in gold metal is 5.90×10^{22} cm^{-3}. Gold is highly conductive to electricity, and has been used for electrical wiring in some high energy applications (silver is even more conductive per volume, but gold has the advantage of corrosion resistance). For example, gold electrical wires were used during some of the Manhattan Project's atomic experiments, but large high current silver wires were used in the calutron isotope separator magnets in the project.



The 220 kg Gold brick displayed in Chinkuashi Gold Museum, Taiwan.

  - Though gold is attacked by free chlorine, its good conductivity and general resistance to oxidation and corrosion in other environments (including resistance to non-chlorinated acids) has led to its widespread industrial use in the electronic era as a thin layer coating electrical connectors of all kinds, thereby ensuring good connection. For example, gold is used in the connectors of the more expensive electronics cables, such as audio, video and USB cables. The benefit of using gold over other connector metals such as tin in these applications, is highly debated. Gold connectors are often criticized by audio-visual experts as unnecessary for most consumers and seen as simply a marketing ploy. However, the use of gold in other applications in electronic sliding contacts in highly humid or corrosive atmospheres, and in use for contacts with a very high failure cost (certain computers, communications equipment, spacecraft, jet aircraft engines) remains very common, and is unlikely to be replaced in the near future by any other metal.
  - Besides sliding electrical contacts, gold is also used in electrical contacts because of its resistance to corrosion, electrical conductivity, ductility and lack of toxicity. Switch contacts are generally subjected to more intense corrosion stress than are sliding contacts.
- Colloidal gold (Colloidal sols of gold nanoparticles) in water are intensely red-colored, and can be made with tightly-controlled particle sizes up to a few tens of nm across by reduction of gold chloride with citrate or ascorbate ions. Colloidal gold is used in research applications in medicine, biology and materials science. The technique of immunogold labeling exploits the ability of the gold particles to adsorb protein molecules onto their surfaces. Colloidal gold particles coated with specific antibodies can be used as probes for the presence and position of antigens on the surfaces of cells (Faulk and Taylor 1979). In ultrathin sections of tissues viewed by electron microscopy, the immunogold labels appear as extremely dense round spots at the position of the antigen (Roth et al. 1980). Colloidal gold is also the form of gold used as gold paint on ceramics prior to firing.
- Gold, or alloys of gold and palladium, are applied as conductive coating to biological specimens and other non-conducting materials such as plastics and glass to be viewed in a scanning electron microscope. The coating, which is usually applied by sputtering with an argon plasma, has a triple role in this application. Gold's very high electrical conductivity drains electrical charge to earth, and its very high density provides stopping power for electrons in the SEM's electron beam, helping to limit the depth to which the electron beam penetrates the specimen. This improves definition of the position and topography of the specimen surface and increases the spatial resolution of the image. Gold also produces a high output of secondary electrons when irradiated by an electron beam, and these low-energy electrons are the most commonly-used signal source used in the scanning electron microscope.
- Many competitions, and honours, such as the Olympics and the Nobel Prize, award a gold medal to the winner.
- As gold is a good reflector of electromagnetic radiation such as infrared and visible light as well as radio waves, it is used for the protective coatings on many artificial satellites, in infrared protective faceplates in thermal protection suits and astronauts' helmets and in electronic warfare planes like the EA-6B Prowler.
- Gold is used as the reflective layer on some high-end CDs.
- The isotope gold-198, (half-life: 2.7 days) is used in some cancer treatments and for treating other diseases.
- Automobiles may use gold for heat insulation. McLaren F1 uses gold foil in the engine compartment.

As gold chemical compounds

Gold is attacked by and dissolves in alkaline solutions of potassium or sodium cyanide, and gold cyanide is the electrolyte used in commercial electroplating of gold onto base metals and electroforming. Gold chloride (chloroauric acid) solutions are used to make colloidal gold by reduction with citrate or ascorbate ions. Gold chloride and gold oxide are used to make highly-valued cranberry or red-colored glass, which, like colloidal gold sols, contains evenly-sized spherical gold nanoparticles.

History


Funerary mask of Tutankhamun

Gold has been known and highly-valued since prehistoric times. It may have been the first metal used by humans and was valued for ornamentation and rituals. Egyptian hieroglyphs from as early as 2600 BC describe gold, which king Tushratta of the Mitanni claimed was "more plentiful than dirt" in Egypt. Egypt and Nubia had the resources to make them major gold-producing areas for much of history. Large mines also occurred across the Red Sea in what is now Saudi Arabia. Gold is also mentioned frequently in the Old Testament, starting with Genesis 2:11 (at Havilah) and is included with the gifts of the magi in the first chapters of Matthew New Testament. The Book of Revelation 21:21 describes the city of New Jerusalem as having streets "made of pure gold, clear as crystal". The south-east corner of the Black Sea was famed for its gold. Exploitation is said to date from the time of Midas, and this gold was important in the establishment of what is probably the world's earliest coinage in Lydia between 643 and 630 BC.

The Romans developed new methods for extracting gold on a large scale using hydraulic mining methods, especially in Spain from 25 BC onwards and in Roumania from 150 AD onwards. One of their largest mines was at Las Medulas in Galicia, where seven long aqueducts enabled them to sluice most of a large alluvial deposit. The mines at Verespatak in Transylvania were also very large, and till very recently, still mined by opencast methods. They also exploited smaller deposits in Wales, such as placer and hard-rock deposits at Dolaucothi. The various methods they used are well described by Pliny the Elder in his encyclopedia Naturalis Historia written towards the end of the first century AD.

The Mali Empire in Africa was famed throughout the old world for its large amounts of gold. Mansa Musa, ruler of the empire (1312–1337) became famous throughout the old world for his great hajj to Mecca in 1324. When he passed through Cairo in July of 1324, he was reportedly accompanied by a camel train that included thousands of people and nearly a hundred camels. He gave away so much gold that it took over a decade for the economy across North Africa to recover, due to the rapid inflation that it initiated. A contemporary Arab historian remarked;

> “ *Gold was at a high price in Egypt until they came in that year. The mithqal did not go below 25 dirhams and was generally above, but from that time its value fell and it cheapened in price and has remained cheap till now. The mithqal does not exceed 22 dirhams or less. This has been the state of affairs for about twelve years until this day by reason of the large amount of gold which they brought into Egypt and spent there [...]* ”
>
> — Chihab Al-Umari

The European exploration of the Americas was fueled in no small part by reports of the gold ornaments displayed in great profusion by Native American peoples, especially in Central America, Peru, and Colombia.

Although the price of some platinum group metals can be much higher, gold has long been considered the most desirable of precious metals, and its value has been used as the standard for many currencies (known as the gold standard) in history. Gold has been used as a symbol for purity, value, royalty, and particularly roles that combine these properties. Gold as a sign of wealth and prestige was made fun of by Thomas More in his treatise *Utopia*. On that imaginary island, gold is so abundant that it is used to make chains for slaves, tableware and lavatory-seats. When ambassadors from other countries arrive, dressed in ostentatious gold jewels and badges, the Utopians mistake them for menial servants, paying homage instead to the most modestly-dressed of their party.

There is an age-old tradition of biting gold in order to test its authenticity. Although this is certainly not a professional way of examining gold, the *bite test* should score the gold because gold is considered a soft metal according to the Mohs' scale of mineral hardness. The purer the gold the easier it should be to mark it. Painted lead can cheat this test because lead is softer than gold (and may invite a small risk of lead poisoning if sufficient lead is absorbed by the biting).

Gold in antiquity was relatively easy to obtain geologically; however, 75% of all gold ever produced has been extracted since 1910. It has been estimated that all the gold in the world that has ever been refined would form a single cube 20 m (66 ft) on a side (equivalent to 8000 m³).

This 156 ounce (4,42 kg) nugget was found by an individual prospector in the Southern California Desert using a metal detector.

One main goal of the alchemists was to produce gold from other substances, such as lead — presumably by the interaction with a mythical substance called the philosopher's stone. Although they never succeeded in this attempt, the alchemists promoted an interest in what can be done with substances, and this laid a foundation for today's chemistry. Their symbol for gold was the circle with a point at its centre (☉), which was also the astrological symbol, the Egyptian hieroglyph and the ancient Chinese character for the Sun. For modern attempts to produce artificial gold, see gold synthesis.

During the 19th century, gold rushes occurred whenever large gold deposits were discovered. The first major gold strike in the United States occurred in a small north Georgia town called Dahlonega. Further gold rushes occurred in California, Colorado, Otago, Australia, Witwatersrand, Black Hills, and Klondike.

Because of its historically high value, much of the gold mined throughout history is still in circulation in one form or another.

Occurrence

In nature, gold most often occurs in its native state (that is, as a metal), though usually alloyed with silver. Native gold contains usually eight to ten percent silver, but often much more — alloys with a silver content over 20% are called electrum. As the amount of silver increases, the colour becomes whiter and the specific gravity becomes lower.

Ores bearing native gold consist of grains or microscopic particles of metallic gold embedded in rock, often in association with veins of quartz or sulfide minerals like pyrite. These are called "lode" deposits. Native gold is also found in the form of free flakes, grains or larger nuggets that have been eroded from rocks and end up in alluvial deposits (called placer deposits). Such free gold is always richer at the surface of gold-bearing veins owing to the oxidation of accompanying minerals followed by weathering, and washing of the dust into streams and rivers, where it collects and can be welded by water action to form nuggets.

Gold sometimes occurs in minerals in chemical composition with other elements, especially in association with tellurium. Examples are calaverite, sylvanite, nagyagite, petzite and krennerite. Gold also occurs rarely as a mercury-gold amalgam, and in very low concentrations in seawater.

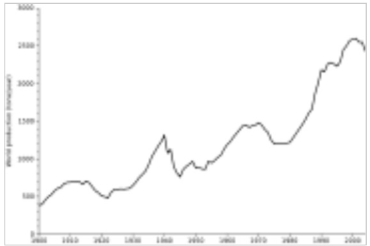
Production


Gold Nuggets found in Arizona


The entrance to an underground gold mine in Victoria, Australia

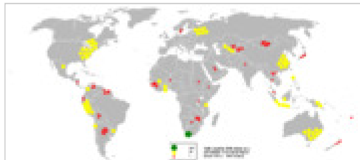
Economic gold extraction can be achieved from ore grades as little as 0.5 g/1000 kg (0.5 parts per million, ppm) on average in large easily mined deposits. Typical ore grades in open-pit mines are 1–5 g/1000 kg (1–5 ppm), ore grades in underground or hard rock mines are usually at least 3 g/1000 kg (3 ppm) on average. Since ore grades of 30 g/1000 kg (30 ppm) are usually needed before gold is visible to the naked eye, in most gold mines the gold is invisible.

Since the 1880s, South Africa has been the source for a large proportion of the world’s gold supply, with about 50% of all gold ever produced having come from South Africa. Production in 1970 accounted for 79% of the world supply, producing about 1,000 tonnes. However by 2007 production was just 272 tonnes. This sharp decline was due to the increasing difficulty of extraction, changing economic factors affecting the industry, and tightened safety auditing. In 2007 China (with 276 tonnes) overtook South Africa as the world's largest gold producer, the first time since 1905 that South Africa has not been the largest.


World gold production trend


Gold ore

The city of Johannesburg located in South Africa was founded as a result of the Witwatersrand Gold Rush which resulted in the discovery of some of the largest gold deposits the world has ever seen. Gold fields located within the basin in the Free State and Gauteng provinces are extensive in strike and dip requiring some of the world's deepest mines, with the Savuka and TauTona mines being currently the world's deepest gold mine at 3,777 m. The Second Boer War of 1899–1901 between the British Empire and the Afrikaner Boers was at least partly over the rights of miners and possession of the gold wealth in South Africa.


Gold output in 2005

Other major producers are United States, Australia, China, Russia and Peru. Mines in South Dakota and Nevada supply two-thirds of gold used in the United States. In South America, the controversial project Pascua Lama aims at exploitation of rich fields in the high mountains of Atacama Desert, at the border between Chile and Argentina. Today about one-quarter of the world gold output is estimated to originate from artisanal or small scale mining.

After initial production, gold is often subsequently refined industrially by the Wohlwill process or the Miller process. Other methods of assaying and purifying smaller amounts of gold include parting and inquartation as well as cuppelation, or refining methods based on the dissolution of gold in aqua regia.

The world's oceans hold a vast amount of gold, but in very low concentrations (perhaps 1–2 parts per 10 billion). A number of people have claimed to be able to economically recover gold from sea water, but so far they have all been either mistaken or crooks. Reverend Prescott Jernegan ran a gold-from seawater swindle in America in the 1890s. A British fraud ran the same scam in England in the early 1900s.

Fritz Haber (the German inventor of the Haber process) attempted commercial extraction of gold from sea water in an effort to help pay Germany's reparations following the First World War. Unfortunately, his assessment of the concentration of gold in sea water was unduly high, probably due to sample contamination. The effort produced little gold and cost the German government far more than the commercial value of the gold recovered. No commercially viable mechanism for performing gold extraction from sea water has yet been identified. Gold synthesis is not economically viable and is unlikely to become so in the foreseeable future.

The average gold mining and extraction costs are $238 per troy ounce but these can vary widely depending on mining type and ore quality. In 2001, global mine production amounted to 2,604 tonnes, or 67% of total gold demand in that year. At the end of 2001, it was estimated that all the gold ever mined totaled 145,000 tonnes.

At current consumption rates, the supply of gold is believed to last 45 years.

Price


LBMA USD morning price fixings ($US per troy ounce) since 2001

Like other precious metals, gold is measured by troy weight and by grams. When it is alloyed with other metals the term *carat* or *karat* is used to indicate the amount of gold present, with 24 karats being pure gold and lower ratings proportionally less. The purity of a gold bar can also be expressed as a decimal figure ranging from 0 to 1, known as the millesimal fineness, such as 0.995 being very pure.

The price of gold is determined on the open market, but a procedure known as the Gold Fixing in London, originating in September 1919, provides a daily benchmark figure to the industry. The afternoon fixing appeared in 1968 to fix a price when US markets are open.

The high price of gold is due to its rare amount. Only three parts out of every billion (0.000000003) in the Earth's crust is gold.

Gold price per ounce in USD since 1968, in actual US$ and 2006 US$

Historically gold was used to back currency; in an economic system known as the gold standard, a certain weight of gold was given the name of a unit of currency. For a long period, the United States government set the value of the US dollar so that one troy ounce was equal to $20.67 ($664.56/kg), but in 1934 the dollar was revalued to $35.00 per troy ounce ($1125.27/kg). By 1961 it was becoming hard to maintain this price, and a pool of US and European banks agreed to manipulate the market to prevent further currency devaluation against increased gold demand.

On 17 March 1968, economic circumstances caused the collapse of the gold pool, and a two-tiered pricing scheme was established whereby gold was still used to settle international accounts at the old $35.00 per troy ounce ($1.13/g) but the price of gold on the private market was allowed to fluctuate; this two-tiered pricing system was abandoned in 1975 when the price of gold was left to find its free-market level. Central banks still hold historical gold reserves as a store of value although the level has generally been declining. The largest gold depository in the world is that of the U.S. Federal Reserve Bank in New York, which holds about 3% of the gold ever mined, as does the similarly-laden U.S. Bullion Depository at Fort Knox.

In 2005 the World Gold Council estimated total global gold supply to be 3,859 tonnes and demand to be 3,754 tonnes, giving a surplus of 105 tonnes.

Price records

Since 1968 the price of gold on the open market has ranged widely, from a high of $850/oz ($27,300/kg) on 21 January 1980, to a low of $252.90/oz ($8,131/kg) on 21 June 1999 (London Gold Fixing). The 1980 high was not overtaken until 3 January 2008 when a new record of $865.35 per troy ounce was set in the a.m. London Gold Fixing. Indexed for inflation, the 1980 high would equate to a price of $2398.21 in 2007 dollars. On January 10, 2008, gold futures for February delivery was $884 on the New York Mercantile Exchange, erasing by $10 the previous record of $875 in 1980, and later it settled at $880.30. On February 26, 2008, gold and platinum dollar prices climbed to historic highs of 948.90 and 2,130.90 dollars per ounce, respectively (London Platinum and Palladium Market). On Thursday, March 13th 2008, gold futures on the New York Mercantile Exchange rose over $1,000/ounce for the first time.

Compounds

Although gold is a noble metal, it forms many and diverse compounds. The oxidation state of gold in its compound ranges from −1 to +5 but Au(I) and Au(III) dominate. Gold(I), referred to as the aurous ion, is the most common oxidation state with “soft” ligands such as thioethers, thiolates, and tertiary phosphines. Au(I) compounds are typically linear. A good example is $Au(CN)_2^-$, which is the soluble form of gold encountered in mining. Curiously, aurous complexes of water are rare. The binary gold halides, such as AuCl, form zig-zag polymeric chains, again featuring linear coordination at Au. Most drugs based on gold are Au(I) derivatives.

Gold(III) (“auric”) is a common oxidation state and is illustrated by gold(III) chloride, $AuCl_3$. Its derivative is chloroauric acid, $HAuCl_4$, which forms when Au dissolves in aqua regia. Au(III) complexes, like other d^8 compounds, are typically square planar.

Less common oxidation states: Au(-I), Au(II), and Au(V)

Compounds containing the Au^- anion are called aurides. Caesium auride, CsAu which crystallizes in the caesium chloride motif. Other aurides include those of Rb^+, K^+, and tetramethylammonium $(CH_3)_4N^+$. Gold(II) compounds are usually diamagnetic with Au-Au bonds such as $[Au(CH_2)_2P(C_6H_5)_2]_2Cl_2$. A noteworthy, legitimate Au(II) complex contains xenon as a ligand, $[AuXe_4](Sb_2F_{11})_2$. Gold pentafluoride is the sole example of Au(V), the highest verified oxidation state.

Some gold compounds exhibit *aurophilic bonding*, which describes the tendency of gold ions to interact at distances that are too long to be a conventional Au-Au bond but shorter that van der Waals bonding. The interaction is estimated to be comparable in strength to that of a hydrogen bond.

Mixed valence compounds

Well-defined cluster compounds are numerous. In such cases, gold has a fractional oxidation state. A representative example is the octahedral species $\{Au(P(C_6H_5)_3)\}_6^{2+}$. Gold chalcogenides, e.g. "AuS" feature equal amounts of Au(I) and Au(III).

Isotopes

There is one stable isotope of gold, and 18 radioisotopes with ^{195}Au being the most stable with a half-life of 186 days.

Gold has been proposed as a " salting" material for nuclear weapons (cobalt is another, better-known salting material). A jacket of natural gold, irradiated by the intense high-energy neutron flux from an exploding thermonuclear weapon, would transmute into the radioactive isotope Au-198 with a half-life of 2.697 days and produce approximately .411 MeV of gamma radiation, significantly increasing the radioactivity of the weapon's fallout for several days. Such a weapon is not known to have ever been built, tested, or used.

Symbolism


Three Gold Sovereigns with a Krugerrand

Gold has been associated with the extremities of utmost evil and great sanctity throughout history. In the Book of Exodus, the Golden Calf is a symbol of idolatry and rebellion against God. In Communist propaganda, the golden pocket watch and its fastening golden chain were the characteristic accessories of the class enemy, the bourgeois and the industrial tycoons. Credit card companies associate their product with wealth by naming and coloring their top-of-the-range cards "gold;" although, in an attempt to out-do each other, platinum (and the even-more-elite black card) has now overtaken gold.

On the other hand in the Book of Genesis, Abraham was said to be rich in gold and silver, and Moses was instructed to cover the Mercy Seat of the Ark of the Covenant with pure gold. Eminent orators such as John Chrysostom were said to have a "mouth of gold with a silver tongue." Gold is associated with notable anniversaries, particularly in a 50-year cycle, such as a golden wedding anniversary, golden jubilee, etc.


Swiss-cast 1 kg gold bar.

Great human achievements are frequently rewarded with gold, in the form of medals and decorations. Winners of races and prizes are usually awarded the gold medal (such as the Olympic Games and the Nobel Prize), while many award statues are depicted in gold (such as the Academy Awards, the Golden Globe Awards the Emmy Awards, the Palme d'Or, and the British Academy Film Awards).

Medieval kings were inaugurated under the signs of sacred oil and a golden crown, the latter symbolizing the eternal shining light of heaven and thus a Christian king's divinely inspired authority. Wedding rings are traditionally made of gold; since it is long-lasting and unaffected by the passage of time, it is considered a suitable material for everyday wear as well as a metaphor for the relationship. In Orthodox Christianity, the wedded couple is adorned with a golden crown during the ceremony, an amalgamation of symbolic rites.

The symbolic value of gold varies greatly around the world, even within geographic regions. For example, gold is quite common in Turkey but considered a most valuable gift in Sicily.

Toxicity

Pure gold is non-toxic and non-irritating when ingested and is sometimes used as a food decoration in the form of gold leaf. It is also a component of the alcoholic drinks Goldschläger, Gold Strike, and Goldwasser. Gold is approved as a food additive in the EU (E175 in the Codex Alimentarius).

Soluble compounds (gold salts) such as potassium gold cyanide, used in gold electroplating, are toxic to the liver and kidneys. There are rare cases of lethal gold poisoning from potassium gold cyanide. Gold toxicity can be ameliorated with chelating agents such as British anti-Lewisite.

Helicopter

2008/9 Schools Wikipedia Selection. Related subjects: Air & Sea transport

Helicopter

A Canadian Helicopters Bell 206

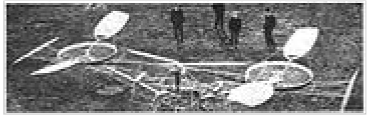
A **helicopter** is an aircraft which is lifted and propelled by one or more horizontal rotors, each rotor consisting of two or more rotor blades. Helicopters are classified as rotorcraft or rotary-wing aircraft to distinguish them from fixed-wing aircraft because the helicopter derives its source of lift from the rotor blades rotating around a mast. The word 'helicopter' is adapted from the French *hélicoptère*, coined by Gustave de Ponton d'Amecourt in 1861. It is linked to the Greek words *helix/helik-* (ἕλικ-) = "spiral" or "turning" and *pteron* (πτερόν) = "wing".

As an aircraft, the primary advantages of the helicopter are due to the rotor blades that revolve through the air, providing lift without requiring the aircraft to move forward the way an airplane does. This creates the ability for the helicopter to take off and land vertically without the need for runways. For this reason, helicopters are often used to operate in congested or isolated areas where airplanes are generally not able to take off or land. The lift from the rotor also allows the helicopter to hover in one area for extended periods of time, and to do so more efficiently than other forms of vertical take-off and landing (VTOL) aircraft, allowing it to accomplish tasks that airplanes are unable to perform.

Although helicopters were developed and built during the first half-century of flight, some even reaching limited production, it wasn't until 1942 that a helicopter designed by Igor Sikorsky became the first helicopter to enter full-scale production, with 131 aircraft built. Even though most previous designs utilized more than one main rotor, it was the single main rotor with antitorque tail rotor configuration of this design that would come to be recognized worldwide as *the helicopter*.

History

Since 400 BC, Chinese children have played with a bamboo flying top. Eventually, this flying top made it to Europe and is depicted in a 1463 European painting. *Pao Phu Tau* (抱朴子) was a 4th-century book in China reported to describe some of the ideas inherent to rotary wing aircraft:


Paul Cornu's helicopter in 1907

> “ Someone asked the master about the principles of mounting to dangerous heights and traveling into the vast inane. The Master said, "Some have made flying cars with wood from the inner part of the jujube tree, using ox-leather [straps] fastened to returning blades so as to set the machine in motion." ”

Leonardo da Vinci conceived a machine that could be described as an "aerial screw". He wrote that he made small flying models but could not stop the rotor from making the whole craft rotate. Later machines would more closely resemble the ancient bamboo flying top, with spinning wings rather than screws.

In July 1754, Mikhail Lomonosov showed the Russian Academy of Sciences a small coaxial rotor powered by a wound-up spring, intended to lift meteorological instruments.

In 1783, Christian de Launoy, and his mechanic, Bienvenu, made a model pair of counter-rotating rotors (not coaxial) using turkey's flight feathers as rotor blades, and in 1784 demonstrated it to the French Academy of Sciences.

In 1861, the word "helicopter" was coined by Gustave de Ponton d'Amécourt, a French inventor who demonstrated a small steam-powered model.

From 1860 to 1880, many small helicopter models were designed and made. These included Alphonse Pénaud's model coaxial rotors, powered by twisted rubber bands (1870). Enrico Forlanini's unmanned helicopter was powered by a steam engine. It was the first of its type that rose to a height of 13 meters, where it remained for some 20 seconds, after a vertical take-off from a park in Milan (1877). Emmanuel Dieuaide's design featured counter-rotating rotors and was steam-powered through a hose from a boiler on the ground (1877). Melikoff designed a man-carrier, but it was almost certainly not built (1877). Dandrieux's design had counter-rotating rotors and a 7.7-pound (3.5-kilogram) steam engine. It rose more than 40 feet (12 meters) and flew for 20 seconds (circa 1878).

In the 1880s, Thomas Edison experimented with small helicopter models in the USA. First with a guncotton-powered engine, which caused damage by explosions, and tests were ended. Next he used an electric motor. His tests showed that a large rotor with low blade area was needed.

Ján Bahýľ, a Slovak inventor, developed a model helicopter powered by an internal combustion engine, that in 1901 reached a height of 0.5 meters. On May 5, 1905 his helicopter reached 4 meters in altitude and flew for over 1500 meters.

First flights

In 1906, two French brothers, Jacques and Louis Breguet, began experimenting with airfoils for helicopters and in 1907, those experiments resulted in the *Gyroplane No.1*. Although there is some discrepancy about the dates, sometime between 14 August and 29 September 1907, the Gyroplane No. 1 lifted its pilot up into the air about two feet (0.6 meters) for a minute. However, the Gyroplane No. 1 proved to be extremely unsteady and required a man at each corner of the airframe to hold it steady. For this reason, the flights of the Gyroplane No. 1 are considered to be the first manned flight of a helicopter, but not a free or untethered flight.

That same year, fellow French inventor Paul Cornu designed and built a helicopter that used two 20-foot (6-meter) counter-rotating rotors driven by a 24-hp

(18-kW) Antoinette engine. On November 13, 1907, it lifted its inventor to 1 foot (0.3 meters) and remained aloft for 20 seconds. Although this flight was smaller in its achievement than that of the Breguet brothers, it was greater in accomplishment in that it was the first true free flight with a pilot. The Cornu helicopter would achieve a height of nearly 2 meters but also proved to be unstable and was abandoned after only a few flights.

Early development

In the early 1920s, Raul Pateras Pescara, an Argentinian working in Europe, demonstrated one of the first successful applications of cyclic pitch. His coaxial, contra-rotating, biplane rotors were able to be warped to cyclically increase and decrease the lift they produced; and the rotor hub could also tilt, both allowing the aircraft to move laterally without a separate propeller to push or pull it. Pescara is also credited with demonstrating the principle of autorotation, the method by which helicopters land safely after engine failure. By January 1924, Pescara's helicopter No. 3 was capable of flights up to 10 minutes.

One of Pescara's contemporaries, Frenchman Etienne Oemichen, set the first helicopter world record recognized by the Fédération Aéronautique Internationale (FAI) on 14 April 1924, flying his helicopter 360 meters (1,181 feet). On 18 April 1924, Pescara beat Oemichen's record, flying for a distance of 736 m (nearly a half mile) in 4 minutes and 11 seconds (about 8 mph, 13 km/h) maintaining a height of six feet. Not to be outdone, Oemichen reclaimed the world record on 4 May when he flew his No. 2 machine again for a 14-minute flight covering 5,550 feet (1.05 mi, 1.692 km) while climbing to a height of 50 feet (15 meters). Oemichen also set the 1-km closed-circuit record at 7 minutes 40 seconds.

Meanwhile, Juan de la Cierva was developing and introducing the first practical rotorcraft in Spain. In 1923, the aircraft that would become the basis for the modern helicopter rotor began to take shape in the form of an autogyro, Cierva's C.4. Cierva had discovered aerodynamic and structural deficiencies in his early designs that could cause his autogyros to flip over after takeoff. The flapping hinges that Cierva designed for the C.4 allowed the rotor to develop lift equally on the left and right halves of the rotor disk. A crash in 1927 led to the development of a drag hinge to relieve further stress on the rotor from its flapping motion. These two developments allowed for a stable rotor system, not only in a hover, but in forward flight.

Albert Gillis von Baumhauer, a Dutch aeronautical engineer, began studying rotorcraft design in 1923. His first prototype 'flew' ('hopped' and hovered really) on September 24, 1925, with Dutch Army-Air arm Captain Floris Albert van Heijst at the controls. The controls that Captain van Heijst used were Von Baumhauer's inventions, the cyclic and collective. Patents were granted to von Baumhauer for his cyclic and collective controls by the British ministry of aviation on 31 January 1927, under patent number 265,272.

In 1930, the Italian engineer Corradino D'Ascanio built his D'AT3, a coaxial helicopter. His relatively large machine had two, two-bladed, counter-rotating rotors. Control was achieved by using auxiliary wings or servo-tabs on the trailing edges of the blades, a concept that was later adopted by other helicopter designers, including Bleeker and Kaman. Three small propellers mounted to the airframe were used for additional pitch, roll, and yaw control. The D'AT3 held modest FAI speed and altitude records for the time, including altitude (18 m), duration (8 minutes 45 seconds) and distance flown (1,078 m).

The Bréguet-Dorand Gyroplane Laboratoire was built in 1933. After many ground tests and an accident, it first took flight on 26 June 1935. Within a short time, the aircraft was setting records with pilot Maurice Claisse at the controls. On 14 December 1935 he set a record for closed-circuit flight with a 500 m diameter. The next year, on 26 September 1936, Claisse set a height record of 158 m. And, finally, on 24 November 1936, he set a flight duration record of one hour, two minutes and 5 seconds over a 44 km closed circuit at 44.7 km/h. The aircraft was destroyed in 1943 by an Allied air strike at Villacoublay airport.

Birth of an industry

Despite the success of the Gyroplane Laboratoire, the German Focke-Wulf Fw 61, first flown in 1936, would eclipse its accomplishments. The Fw 61 broke all of the helicopter world records in 1937, demonstrating a flight envelope that had only previously been achieved by the autogyro. In February 1938, Hanna Reitsch became the first female helicopter pilot, exhibiting the Fw 61 before crowds in the Deutschlandhalle.

Nazi Germany would use helicopters in small numbers during World War II for observation, transport, and medical evacuation. The Flettner Fl 282 *Kolibri* synchropter was used in the Mediterranean Sea, while the Focke Achgelis Fa 223 *Drache* was used in Europe. Extensive bombing by the Allied forces prevented Germany from producing any helicopters in large quantities during the war.

In the United States, Igor Sikorsky and W. Lawrence LePage, were competing to produce the United States military's first helicopter. Prior to the war, LePage had received the patent rights to develop helicopters patterned after the Fw 61, and built the XR-1, utilizing the transverse rotor layout. Meanwhile, Sikorsky, had settled on a simpler, single rotor design, the VS-300. After experimenting with configurations to counteract the torque produced by the single main rotor, he settled on a single, smaller rotor mounted vertically on the tailboom.

Developed from the VS-300, Sikorsky's R-4 became the first mass produced helicopter with a production order for 100 aircraft. The R-4 was the only Allied helicopter to see service in World War II, primarily being used for rescue in Burma, Alaska, and other areas with harsh terrain. Total production would reach 131 helicopters before the R-4 was replaced by other Sikorsky helicopters such as the R-5 and the R-6. In all, Sikorsky would produce over 400 helicopters before the end of World War II.

As LePage and Sikorsky were building their helicopters for the military, Bell Aircraft hired Arthur Young to help build a helicopter using Young's semi-rigid, teetering-blade rotor design, which utilized a weighted stabilizing bar. The subsequent Model 30 helicopter demonstrated the simplicity and ease of the design. The Model 30 was developed into the Bell 47, which became the first aircraft certificated for civilian use in the United States. Produced in several countries, the Bell 47 would become the most popular helicopter model for nearly 30 years.

Turbine age

In 1951, at the urging of his contacts at the Department of the Navy, Charles H. Kaman modified his Ka-225 helicopter with a new kind of engine, the turboshaft engine. This adaptation of the turbine engine provided a large amount of horsepower to the helicopter with a lower weight penalty than piston engines, with their heavy engine blocks and auxiliary components. On 11 December 1951, the Ka-225 became the first turbine-powered helicopter in the world. Two years later, on 26 March 1954, a modified Navy HTK-1, another Kaman helicopter, became the first twin-turbine helicopter to fly. However, it was the Sud Aviation Alouette II that would become the first helicopter to be produced with a turbine-engine.

Reliable helicopters capable of stable hover flight were developed decades after fixed-wing aircraft. This is largely due to higher engine power density requirements than fixed-wing aircraft. Improvements in fuels and engines during the first half of the 20th century were a critical factor in helicopter development. The availability of lightweight turboshaft engines in the second half of the 20th century led to the development of larger, faster, and higher-performance helicopters. While smaller and less expensive helicopters still use piston engines, turboshaft engines are the preferred powerplant for helicopters today.

Uses

Due to the unique operating characteristics of the helicopter—its ability to takeoff and land vertically, and to hover for extended periods of time, as well as the aircraft's handling properties under low airspeed conditions—it has grown increasingly popular for conducting tasks that were previously not possible, or were time- or work-intensive. Today, helicopters are used for transportation, for construction, for firefighting, search and rescue, and a variety of other jobs that require the special capabilities of the helicopter.

Sikorsky S-64 Skycrane

As aerial cranes, helicopters carry loads connected to long cables or slings in order to place heavy equipment such as transmission towers and large air conditioning units on the tops of tall buildings or when an item must be raised up in a remote area, such as a radio tower raised on the top of a hill or mountain, far from the nearest road. The most popular use of helicopters as aerial cranes is in the logging industry to lift large trees out of rugged terrain where vehicles aren't able to reach, or where environmental concerns prohibit the building of roads. These operations are referred to as longline because of the long, single sling line used to carry the load.

Kern County (California) Fire Department Bell 205 dropping water on fire

Aerial firefighting (or water bombing) is a method to combat wildfires that often uses helicopters. Helicopters may be fitted with tanks or carry buckets or deliver firefighters who rappel to the ground below. Buckets, such as the Bambi bucket, are usually filled by submerging in lakes, rivers, reservoirs, or portable tanks. Tanks may be filled on the ground or water may be siphoned from lakes or reservoirs through a hanging snorkel. Helicopters are also used to resupply firefighters on the ground with tools, food, water and other supplies. Popular firefighting helicopters include variants of the Bell 205 and the Erickson S-64 Aircrane helitanker.

Helicopters are used as an air ambulance for emergency medical assistance in situations where either a traditional ambulance cannot easily or quickly reach the scene or when a patient needs to be transported at a distance where air transportation is most practical. Air ambulance helicopters are equipped to provide medical treatment to a critically injured or ill patient while in flight. The use of helicopters as an air ambulance is often referred to as MEDEVAC, and patients are referred to as being "airlifted", or "medevaced".

Polish police Bell 206

Police departments and other law enforcement agencies use helicopters to search for and pursue suspects. Since helicopters can achieve a unique aerial view and don't need to negotiate ground obstacles, they are often used in conjunction with police on the ground to report on suspects' locations and movements. They are often mounted with lighting and heat-sensing equipment for night pursuits.

Military forces use attack helicopters to conduct aerial attacks on ground targets. Such helicopters are mounted with missile launchers and miniguns. Transport helicopters are used to ferry troops and supplies where the lack of an airstrip would make transport via fixed-wing aircraft impossible. Transport helicopters used to deliver troops as an attack force on an objective is referred to as Air Assault.

Other uses

- Aerial photography
- Motion picture photography
- Electronic news gathering
- Search and Rescue
- Touring or personal pleasure
- Transport

Rotor configurations

Most helicopters have a single, main rotor but require a separate rotor to overcome torque. This is accomplished through a variable pitch, antitorque rotor or tail rotor. This is the design that Igor Sikorsky settled on for his VS-300 helicopter and it has become the recognized convention for helicopter design, although, designs do vary. When viewed from above, designs from Germany, United Kingdom and the United States are said to rotate counter-clockwise, all others are said to rotate clockwise. This can make it difficult when discussing aerodynamic effects on the main rotor between different designs, since the effects may manifest on opposite sides of each aircraft.

Single main rotor

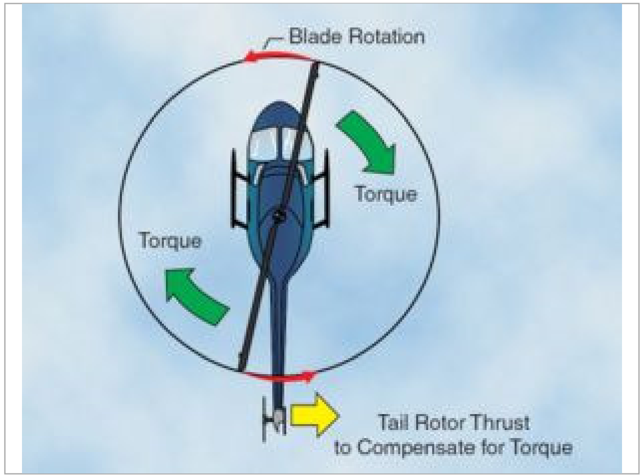
With a single main rotor helicopter, the creation of torque as the engine turns the rotor creates a torque effect which causes the body of the helicopter to turn in the opposite direction of the rotor. To eliminate this effect, some sort of antitorque control must be used, with a sufficient margin of power available to allow the helicopter to maintain its heading and provide yaw control. The three most common controls used today are the traditional *tail rotor*, Eurocopter's *Fenestron* (also called a *fantail*), and MD Helicopters' *NOTAR*.

Antitorque


Tail rotor of an SA 330 Puma

Tail rotor

The tail rotor is a smaller rotor mounted vertically or near-vertical on the tail of a traditional single-rotor helicopter. The tail rotor either pushes or pulls against the tail to counter the torque. The tail rotor drive system consists of a drive shaft powered from the main transmission and a gearbox mounted at the end of the tail boom. The drive shaft may consist of one long shaft or a series of shorter shafts connected at both ends with flexible couplings. The flexible couplings allow the drive shaft to flex with the tail boom. The gearbox at the end of the tailboom provides an angled drive for the tail rotor and may also include gearing to adjust the output to the optimum RPM for the tail rotor. On some larger helicopters, intermediate gearboxes are used to transition the tail rotor drive shaft from along the tailboom or tailcone to the top of the tail rotor pylon which also serves as a vertical stabilizing airfoil to alleviate the power requirement for the tail rotor in forward flight. It may also serve to provide limited antitorque within certain airspeed ranges in the event that the tail rotor or the tail rotor flight controls fail.



Torque effect on a helicopter

Ducted fan

Fenestron and FANTAIL are trademarks for a ducted fan mounted at the end of the tail boom of the helicopter and used in place of a tail rotor. Ducted fans have between 8 and 18 blades arranged with irregular spacing, so that the noise is distributed over different frequencies. The housing is integral with the aircraft skin and allows a high rotational speed, therefore a ducted fan can have a smaller size than a conventional tail rotor.


Fenestron on a EC 120B

The Fenestron was used for the first time at the end of the 1960s on the second experimental model of Sud Aviation's SA 340, and produced on the later model Aérospatiale SA 341 Gazelle. Besides Eurocopter and its predecessors, a ducted fan tail rotor was also used on the canceled military helicopter project, the United States Army's RAH-66 Comanche, as the FANTAIL.

NOTAR

NOTAR, an acronym for *NO TAil Rotor*, is a relatively new helicopter anti-torque system which eliminates the use of the tail rotor on a helicopter. Although the concept took some time to refine, the NOTAR system is simple in theory and works to provide antitorque the same way a wing develops lift using the Coandă effect. A variable pitch fan is enclosed in the aft fuselage section immediately forward of the tail boom and driven by the main rotor transmission. This fan forces low pressure air through two slots on the right side of the tailboom, causing the downwash from the main rotor to hug the tailboom, producing lift, and thus a measure of antitorque proportional to the amount of airflow from the rotorwash. This is augmented by a direct jet thruster (which also provides directional yaw control) and vertical stabilizers.

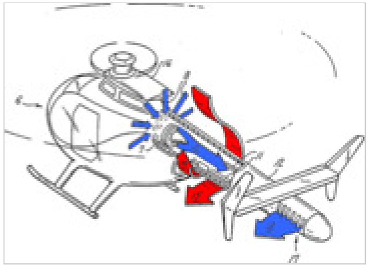


Diagram showing the movement of air through the NOTAR system.

Development of the NOTAR system dates back to 1975 when engineers at Hughes Helicopters began concept development work. In December 1981 Hughes flew a OH-6A fitted with NOTAR for the first time. A more heavily modified prototype demonstrator first flew in March 1986 and successfully completed an advanced flight-test program, validating the system for future application in helicopter design. There are currently three production helicopters that utilize the NOTAR system, all produced by MD Helicopters. This antitorque design also improves safety by eliminating the opportunity for personnel to walk into the tail rotor.

MD Helicopters 520N NOTAR

Tip jets

Another single main rotor configuration without a tail rotor is the tip jet rotor, where the main rotor is not driven by the mast, but from nozzles on the tip of the rotor blade; which are either pressurized from a fuselage-mounted gas turbine or have their own turbojet, ramjet or rocket thrusters. Although this method is simple and eliminates torque, the prototypes that have been built are less fuel efficient than conventional helicopters and produce more noise. One example, the Percival P.74, was not even able to leave the ground, and the Hiller YH-32 Hornet had good lifting capability but was otherwise poor. The Fairey Jet Gyrodyne and 40-seat Fairey Rotodyne flew very well indeed. Possibly the most unusual was the rocket tipped Rotary Rocket Roton ATV. None have made it into production.

Dual rotors (contra-rotating)

Contra-rotating rotors, are rotorcraft configurations with a pair or more of large horizontal rotors turning in opposite directions to counteract the effects of torque on the aircraft without relying on an antitorque tail rotor. Primarily, there are three common configurations that utilize the contra-rotating effect to benefit the rotorcraft; tandem rotors are two rotors with one mounted behind the other, coaxial rotors are two rotors that are mounted one above the other with the same axis, and intermeshing rotors are two rotors that are mounted close to each other at enough angle to allow the rotors to intermesh over the top of the aircraft. Another configuration found on tiltrotors and some earlier helicopters is called transverse rotors where the pair of rotors is mounted at each end of a wing-type structure or outriggers.

Tandem

Tandem rotors are two horizontal main rotor assemblies mounted one behind the other with the rear rotor mounted slightly higher than the front rotor. Tandem rotors achieve pitch attitude changes to accelerate and decelerate the helicopter through a process called differential collective pitch. To pitch forward and accelerate, the rear rotor increases collective pitch, raising the tail and the front rotor decreases collective pitch, simultaneously dipping the nose. To pitch upward while decelerating (or moving rearward), the front rotor increases collective pitch to raise the nose and the rear rotor decreases collective pitch to lower the tail. Yaw control is developed through opposing cyclic pitch in each rotor; to pivot right, the front rotor tilts right and the rear rotor tilts left, and to pivot left, the front rotor tilts left and the rear rotor tilts right.

CH-47 Chinook

Coaxial

Coaxial rotors are a pair of rotors turning in opposite directions, but mounted on a mast, with the same axis of rotation, one above the other. The advantage of the coaxial rotor is that, in forward flight, the lift provided by the advancing halves of each rotor compensates for the retreating half of the other, eliminating one of the key effects of dissymmetry of lift; retreating blade stall. However, other design considerations plague coaxial rotors. There is an increased mechanical complexity of the rotor system because it requires linkages and swashplates for two rotor systems. Add that each rotor system needs to be turned in opposite directions means that the mast itself is more complex, and provisions for making pitch changes to the upper rotor system must pass through the lower rotor system.

Kamov Ka-50

Intermeshing

Intermeshing rotors on a helicopter are a set of two rotors turning in opposite directions, with each rotor mast mounted on the helicopter with a slight angle to the other so that the blades intermesh without colliding. This configuration is sometimes referred to as a synchropter. Intermeshing rotors have high stability and powerful lifting capability. The arrangement was successfully used in Nazi Germany for a small anti-submarine warfare helicopter, the Flettner Fl 282 Kolibri. During the Cold War, the American company, Kaman Aircraft produced the HH-43 Huskie for the USAF firefighting and rescue missions. The latest Kaman model, the Kaman K-MAX, is a dedicated sky crane design.

HH-43 Huskie

Transverse


Mi-12

Transverse rotors are mounted on the end of wings or outriggers, perpendicular to the body of the aircraft. Similar to tandem rotors and intermeshing rotors, the transverse rotor also utilizes differential collective pitch. But like the intermeshing rotors, the transverse rotors use the concept for changes in the roll attitude of the rotorcraft. This configuration is found on two of the first viable helicopters, the Focke-Wulf Fw 61 and the Focke-Achgelis Fa 223, as well as the world's largest helicopter ever built, the Mil Mi-12. It is also the configuration found on tiltrotors, such as Bell's XV-15 and the newer V-22 Osprey.

Helicopter rotor system

The rotor system, or more simply *rotor*, is the rotating part of a helicopter which generates lift. A rotor system may be mounted horizontally as main rotors are, providing lift vertically, or it may be mounted vertically, such as a tail rotor, to provide lift horizontally as thrust to counteract torque effect. In the case of tiltrotors, the rotor is mounted on a nacelle that rotates at the edge of the wing to transition the rotor from a horizontal mounted position, providing lift horizontally as thrust, to a vertical mounted position providing lift exactly as a helicopter.

The rotor consists of a mast, hub and rotor blades. The *mast* is a cylindrical metal shaft which extends upwards from and is driven by the transmission. At the top of the mast is the attachment point for the rotor blades called the *hub*. The rotor blades are then attached to the hub by a number of different methods. Main rotor systems are classified according to how the main rotor blades are attached and move relative to the main rotor hub. There are three basic classifications: semirigid, rigid, or fully articulated, although some modern rotor systems use an engineered combination of these types.

Semirigid


Semirigid rotor system

A semirigid rotor system allows for two different movements, flapping and feathering. This system is normally comprised of two blades, which are rigidly attached to the rotor hub. The hub is then attached to the rotor mast by a trunnion bearing or teetering hinge and is free to tilt with respect to the main rotor shaft. This allows the blades to see-saw or flap together. As one blade flaps down, the other flaps up. Feathering is accomplished by the feathering hinge, which changes the pitch angle of the blade. Since there is no vertical drag hinge, lead-lag forces are absorbed through blade bending.

Helicopters with semi-rigid rotors are vulnerable to a condition known as mast bumping which can cause the rotor flap stops to shear the mast. Mast bumping is normally encountered during low-G maneuvers, so it is written into the operator's handbook to avoid any low-G conditions.

Fully articulated

In a fully articulated rotor system, each rotor blade is attached to the rotor hub through a series of hinges, which allow the blade to move independently of the others. These rotor systems usually have three or more blades. The blades are allowed to flap, feather, and lead or lag independently of each other. The horizontal hinge, called the flapping hinge, allows the blade to move up and down. This movement is called flapping and is designed to compensate for dissymmetry of lift. The flapping hinge may be located at varying distances from the rotor hub, and there may be more than one hinge. The vertical hinge, called the lead-lag or drag hinge, allows the blade to move back and forth. This movement is called lead-lag, dragging, or hunting. Dampers are usually used to prevent excess back and forth movement around the drag hinge. The purpose of the drag hinge and dampers is to compensate for the acceleration and deceleration caused by Coriolis Effect. Each blade can also be feathered, that is, rotated around its spanwise axis. Feathering the blade means changing the pitch angle of the blade. By changing the pitch angle of the blades you can control the thrust and direction of the main rotor disc.

Rigid

In a rigid rotor system, the blades, hub, and mast are rigid with respect to each other. The rigid rotor system is mechanically simpler than the fully articulated rotor system. There are no vertical or horizontal hinges so the blades cannot flap or drag, but they can be feathered. Operating loads from flapping and lead/lag forces must be absorbed by bending rather than through hinges. By flexing, the blades themselves compensate for the forces which previously required rugged hinges. The result is a rotor system that has less lag in the control response, because the rotor has much less oscillation. The rigid rotor system also negates the danger of mast bumping inherent in semi-rigid rotors.

Combination

Modern rotor systems may use the combined principles of the rotor systems mentioned above. Some rotor hubs incorporate a flexible hub, which allows for blade bending (flexing) without the need for bearings or hinges. These systems, called flextures, are usually constructed from composite material. Elastomeric bearings may also be used in place of conventional roller bearings. Elastomeric bearings are bearings constructed from a rubber type material and have limited movement that is perfectly suited for helicopter applications. Flextures and elastomeric bearings require no lubrication and, therefore, require less maintenance. They also absorb vibration, which means less fatigue and longer service life for the helicopter components.

Controlling flight

A typical helicopter has four separate flight control inputs. These are the cyclic, the collective, the anti-torque pedals, and the throttle. The cyclic control is usually located between the pilot's legs and is commonly called the *cyclic stick* or just *cyclic*. On most helicopters, the cyclic is similar to a joystick. Although, the Robinson R22 and R44 have a unique teetering bar cyclic control system and a few helicopters have a cyclic control that descends into the cockpit from overhead.

Cockpit of an Alouette III

The control is called the cyclic because it changes the pitch of the rotor blades cyclically. The result is to tilt the rotor disk in a particular direction, resulting in the helicopter moving in that direction. If the pilot pushes the cyclic forward, the rotor disk tilts forward, and the rotor produces a thrust in the forward direction. If the pilot pushes the cyclic to the side, the rotor disk tilts to that side and produces thrust in that direction, causing the helicopter to hover sideways.

The collective pitch control or *collective* is located on the left side of the pilot's seat with a settable friction control to prevent inadvertent movement. The collective changes the pitch angle of all the main rotor blades collectively (i.e. all at the same time) and independently of their position. Therefore, if a collective input is made, all the blades change equally, and the result is the helicopter increasing or decreasing in altitude.

The anti-torque pedals are located in the same position as the rudder pedals in an airplane, and serve a similar purpose, namely to control the direction in which the nose of the aircraft is pointed. Application of the pedal in a given direction changes the pitch of the tail rotor blades, increasing or reducing the thrust produced by the tail rotor and causing the nose to yaw in the direction of the applied pedal. The pedals mechanically change the pitch of the tail rotor altering the amount of thrust produced.

Helicopter rotors are designed to operate at a specific RPM. The throttle controls the power produced by the engine, which is connected to the rotor by a transmission. The purpose of the throttle is to maintain enough engine power to keep the rotor RPM within allowable limits in order to keep the rotor producing enough lift for flight. In single-engine helicopters, the throttle control is a motorcycle-style twist grip mounted on the collective control, while dual-engine helicopters have a power lever for each engine.

Flight conditions

There are two basic flight conditions for a helicopter; hover and forward flight.

- Hover

  Hovering is the most challenging part of flying a helicopter. This is because a helicopter generates its own gusty air while in a hover, which acts against the fuselage and flight control surfaces. The end result is constant control inputs and corrections by the pilot to keep the helicopter where it is required to be. Despite the complexity of the task, the control inputs in a hover are simple. The cyclic is used to eliminate drift in the horizontal plane, that is to control forward and back, right and left. The collective is used to maintain altitude. The pedals are used to control nose direction or heading. It is the interaction of these controls that makes hovering so difficult, since an adjustment in any one control requires an adjustment of the other two, creating a cycle of constant correction.

- Forward flight

  In forward flight a helicopter's flight controls behave more like that in a fixed-wing aircraft. Displacing the cyclic forward will cause the nose to pitch down, with a resultant increase in airspeed and loss of altitude. Aft cyclic will cause the nose to pitch up, slowing the helicopter and causing it to climb. Increasing collective (power) while maintaining a constant airspeed will induce a climb while decreasing collective will cause a descent. Coordinating these two inputs, down collective plus aft cyclic or up collective plus forward cyclic, will result in airspeed changes while maintaining a constant altitude. The pedals serve the same function in both a helicopter and an airplane, to maintain balanced flight. This is done by applying a pedal input in whichever direction is necessary to centre the ball in the turn and bank indicator.

Limitations


A-Star pulling rope through skywire traveller

The single most obvious limitation of the helicopter is its slow speed. There are several reasons why a helicopter cannot fly as fast as a fixed wing aircraft. When the helicopter is at rest, the outer tips of the rotor travel at a speed determined by the length of the blade and the RPM. In a moving helicopter, however, the speed of the blades relative to the air depends on the speed of the helicopter as well as on their rotational velocity. The airspeed of the advancing rotor blade is much higher than that of the helicopter itself. It is possible for this blade to exceed the speed of sound, and thus produce vastly increased drag and vibration.

Because the advancing blade has higher airspeed than the retreating blade and generates a dissymmetry of lift, rotor blades are designed to "flap" – lift and twist in such a way that the advancing blade flaps up and develops a smaller angle of attack. Conversely, the retreating blade flaps down, develops a higher angle of attack, and generates more lift. At high speeds, the force on the rotors is such that they "flap" excessively and the retreating blade can reach too high an angle and stall. For this reason, the maximum safe forward speed of a helicopter is given a design rating called V_{NE}, *Velocity, Never Exceed*.


Boeing CH-47 Chinook

During the closing years of the 20th century designers began working on helicopter noise reduction. Urban communities have often expressed great dislike of noisy aircraft, and police and passenger helicopters can be unpopular. The redesigns followed the closure of some city heliports and government action to constrain flight paths in national parks and other places of natural beauty.

Helicopters vibrate. An unadjusted helicopter can easily vibrate so much that it will shake itself apart. To reduce vibration, all helicopters have rotor adjustments for height and pitch. Most also have vibration dampers for height and pitch. Some also use mechanical feedback systems to sense and counter vibration. Usually the feedback system uses a mass as a "stable reference" and a linkage from the mass operates a flap to adjust the rotor's angle of attack to counter the vibration. Adjustment is difficult in part because measurement of the vibration is hard. The most common adjustment measurement system is to use a stroboscopic flash lamp, and observe painted markings or coloured reflectors on the underside of the rotor blades. The traditional low-tech system is to mount coloured chalk on the rotor tips, and see how they mark a linen sheet.


The Bristol Type 192 Belvedere.

Hazards of helicopter flight

As with any moving vehicle, operation outside of safe regimes could result in loss of control, structural damage, or fatality. The following is a list of some of the potential hazards for helicopters:

- Settling with power, also known as a vortex ring state, is essentially when the aircraft settles into its own downwash, unable to climb out of the condition due to the effect of the turbulent air on the aerodynamics of the rotor.
- Retreating blade stall
- Ground resonance (affects helicopters with rotor systems having lead-lag natural frequency less than the blade rotation frequency).
- Low-G condition (affects two-bladed main rotor helicopters)
- Dynamic rollover

- Operating within the shaded area of the height-velocity diagram
- Tail rotor failure.
- Brownout

Deadliest helicopter crashes:

1. Khankala attack - Mi-26 shot down over Chechnya in 2002; 127 dead.
2. 1997 Israeli helicopter disaster - MH-53 crash in Israel in 1997; 73 killed.
3. 1977 Israeli CH-53 crash - CH-53 crash near Yitav-in the Jordan Valley on 10 May 1977; 54 killed.

Jet engine

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

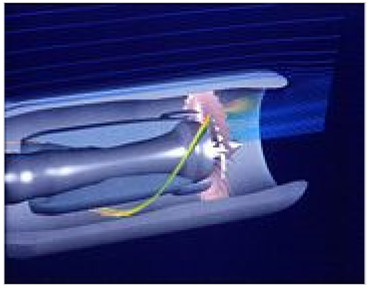

A Pratt & Whitney F100 turbofan engine for the F-15 Eagle and the F-16 Falcon is tested at Robins Air Force Base, Georgia, USA. The tunnel behind the engine muffles noise and allows exhaust to escape

A **jet engine** is a reaction engine that discharges a fast moving jet of fluid to generate thrust in accordance with Newton's third law of motion. This broad definition of jet engines includes turbojets, turbofans, rockets, ramjets, pulse jets and pump-jets. In general, most jet engines are internal combustion engines but non-combusting forms also exist.

In common usage, the term 'jet engine' generally refers to a gas turbine driven internal combustion engine, an engine with a rotary compressor powered by a turbine (" Brayton cycle"), with the leftover power providing thrust. These types of jet engines are primarily used by jet aircraft for long distance travel. The early jet aircraft used turbojet engines which were relatively inefficient for subsonic flight. Modern jet aircraft usually use high-bypass turbofan engines which help give high speeds as well as, over long distances, giving better fuel efficiency than many other forms of transport.

About 7.2% of the oil used in 2004 was ultimately consumed by jet engines In 2007, the cost of jet fuel, while highly variable from one airline to another, averaged 26.5% of total operating costs, making it the single largest operating expense for most airlines.

History


Simulation of a turbojet's airflow

Jet engines can be dated back to the first century AD, when Hero of Alexandria invented the aeolipile. This used steam power directed through two jet nozzles so as to cause a sphere to spin rapidly on its axis. So far as is known, it was little used for supplying mechanical power, and the potential practical applications of Hero's invention of the jet engine were not recognized. It was simply considered a curiosity.

Jet propulsion only literally and figuratively took off with the invention of the rocket by the Chinese in the 11th century. Rocket exhaust was initially used in a modest way for fireworks but gradually progressed to propel formidable weaponry; and there the technology stalled for hundreds of years.

In Ottoman Turkey in 1633 Lagari Hasan Çelebi took off with what was described to be a cone shaped rocket and then glided with wings into a successful landing winning a position in the Ottoman army. However, this was essentially a stunt.

The problem was that rockets are simply too inefficient at low speeds to be useful for general aviation. In 1913 René Lorin came up with a form of jet engine, the subsonic ramjet, which would have been somewhat more efficient, but he had no way to achieve high enough speeds for it to operate, and the concept remained theoretical for quite some time.

However, engineers were beginning to realize that the piston engine was self-limiting in terms of the maximum performance which could be attained; the limit was essentially one of propeller efficiency. This seemed to peak as blade tips approached the speed of sound. If engine, and thus aircraft, performance were ever to increase beyond such a barrier, a way would have to be found to radically improve the design of the piston engine, or a wholly new type of powerplant would have to be developed. This was the motivation behind the development of the gas turbine engine, commonly called a "jet" engine, which would become almost as revolutionary to aviation as the Wright brothers' first flight.

The earliest attempts at jet engines were hybrid designs in which an external power source first compressed air, which was then mixed with fuel and burned for jet thrust. In one such system, called a thermojet by Secondo Campini but more commonly, motorjet, the air was compressed by a fan driven by a conventional piston engine. Examples of this type of design were Henri Coandă's Coandă-1910 aircraft, and the much later Campini Caproni CC.2, and the Japanese Tsu-11 engine intended to power Ohka kamikaze planes towards the end of World War II. None were entirely successful and the CC.2 ended up being slower than the same design with a traditional engine and propeller combination.

The key to a practical jet engine was the gas turbine, used to extract energy from the engine itself to drive the compressor. The gas turbine was not an idea developed in the 1930s: the patent for a stationary turbine was granted to John Barber in England in 1791. The first gas turbine to successfully run self-sustaining was built in 1903 by Norwegian engineer Ægidius Elling. The first patents for jet *propulsion* were issued in 1917. Limitations in design and practical engineering and metallurgy prevented such engines reaching manufacture. The main problems were safety, reliability, weight and, especially, sustained operation.

Albert Fonó In 1915 devised a solution for increasing the range of artillery, comprising a gun-luanched projectile which was to be united with a ramjet propulsion unit. This was to make it possible to obtain a long range with low initial muzzle velocities, allowing heavy shells to be fired from relatively lightweight guns. Fonó submitted his invention to the Austro-Hungarian Army but the proposal was rejected. In 1928 he applied for a German patent on supersonic ramjets, and this was awarded in 1932.

In 1923, Edgar Buckingham of the US National Bureau of Standard published a report expressing scepticism that jet engines would be economically competitive with prop driven aircraft at the low altitudes and airspeeds of the period: "there does not appear to be, at present, any prospect whatever that jet propulsion of the sort here considered will ever be of practical value, even for military purposes."

Instead, by the 1930s, the piston engine in its many different forms (rotary and static radial, aircooled and liquid-cooled inline) was the only type of powerplant available to aircraft designers. This was acceptable as long as only low performance aircraft were required, and indeed all that were available.

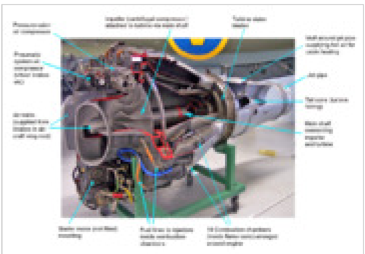

The Whittle W.2/700 engine flew in the Gloster E.28/39, the first British aircraft to fly with a turbojet engine, and the Gloster Meteor

In 1928, RAF College Cranwell cadet Frank Whittle formally submitted his ideas for a turbo-jet to his superiors. In October 1929 he developed his ideas further. . On 16 January 1930 in England, Whittle submitted his first patent (granted in 1932). The patent showed a two-stage axial compressor feeding a single-sided centrifugal compressor. Practical axial compressors were made possible by ideas from A.A.Griffith in a seminal paper in 1926 ("An Aerodynamic Theory of Turbine Design"). Whittle would later concentrate on the simpler centrifugal compressor only, for a variety of practical reasons. Whittle had his first engine running in April 1937. It was liquid-fuelled, and included a self-contained fuel pump. Whittle's team experienced near-panic when the engine would not stop, accelerating even after the fuel was switched off. It turned out that fuel had leaked into the engine and accumulated in pools. So the engine would not stop until all the leaked fuel had burned off. Whittle was unable to interest the government in his invention, and development continued at a slow pace.


Heinkel He 178, the world's first aircraft to fly purely on turbojet power

In 1935 Hans von Ohain started work on a similar design in Germany, apparently unaware of Whittle's work. His first engine was strictly experimental and could only run under external power, but he was able to demonstrate the basic concept. Ohain was then introduced to Ernst Heinkel, one of the larger aircraft industrialists of the day, who immediately saw the promise of the design. Heinkel had recently purchased the Hirth engine company, and Ohain and his master machinist Max Hahn were set up there as a new division of the Hirth company. They had their first HeS 1 centrifugal engine running by September 1937. Unlike Whittle's design, Ohain used hydrogen as fuel, supplied under external pressure. Their subsequent designs culminated in the gasoline-fuelled HeS 3 of 1,100 lbf (5 kN), which was fitted to Heinkel's simple and compact He 178 airframe and flown by Erich Warsitz in the early morning of August 27, 1939, from Marienehe aerodrome, an impressively short time for development. The He 178 was the **world's first jet plane.**

Meanwhile, Whittle's engine was starting to look useful, and his **Power Jets Ltd.** started receiving Air Ministry money. In 1941 a flyable version of the engine called the **W.1**, capable of 1000 lbf (4 kN) of thrust, was fitted to the Gloster E28/39 airframe specially built for it, and first flew on May 15, 1941 at RAF Cranwell.


A picture of an early centrifugal engine (DH Goblin II) sectioned to show its internal components

A British aircraft engine designer, Frank Halford, working from Whittle's ideas developed a "straight through" version of the centrifugal jet; his design became the de Havilland Goblin.

One problem with both of these early designs, which are called **centrifugal-flow** engines, was that the compressor worked by "throwing" (accelerating) air outward from the central intake to the outer periphery of the engine, where the air was then compressed by a divergent duct setup, converting its velocity into pressure. An advantage of this design was that it was already well understood, having been implemented in centrifugal superchargers, then in widespread use on piston engines. However, given the early technological limitations on the shaft speed of the engine, the compressor needed to have a very large diameter to produce the power required. This meant that the engines had a large frontal area, which made it less useful as an aircraft powerplant due to drag. A further disadvantage was that the air flow had to be "bent" to flow rearwards through the combustion section and to the turbine and tailpipe, adding complexity and lowering efficiency. Nevertheless, these types of engines had the major advantages of light weight, simplicity and reliability, and development rapidly progressed to practical airworthy designs.


A cutaway of the Junkers Jumo 004 engine.

Austrian Anselm Franz of Junkers' engine division (*Junkers Motoren* or **Jumo**) addressed these problems with the introduction of the axial-flow compressor. Essentially, this is a turbine in reverse. Air coming in the front of the engine is blown towards the rear of the engine by a fan stage (convergent ducts), where it is crushed against a set of non-rotating blades called *stators* (divergent ducts). The process is nowhere near as powerful as the centrifugal compressor, so a number of these pairs of fans and stators are placed in series to get the needed compression. Even with all the added complexity, the resulting engine is much smaller in diameter and thus, more aerodynamic. Jumo was assigned the next engine number in the RLM numbering sequence, 4, and the result was the Jumo 004 engine. After many lesser technical difficulties were solved, mass production of this engine started in 1944 as a powerplant for the world's first jet-fighter aircraft, the Messerschmitt Me 262 (and later the world's first jet-bomber aircraft, the Arado Ar 234). A variety of reasons conspired to delay the engine's availability, this delay caused the fighter to arrive too late to decisively impact Germany's position in World War II. Nonetheless, it will be remembered as the first use of jet engines in service.

In the UK, their first axial-flow engine, the Metrovick F.2, ran in 1941 and was first flown in 1943. Although more powerful than the centrifugal designs at the time, the Ministry considered its complexity and unreliability a drawback in wartime. The work at Metrovick led to the Armstrong Siddeley Sapphire engine which would be built in the US as the J65.

Following the end of the war the German jet aircraft and jet engines were extensively studied by the victorious allies and contributed to work on early Soviet and US jet fighters. The legacy of the axial-flow engine is seen in the fact that practically all jet engines on fixed wing aircraft have had some inspiration from this design.

Centrifugal-flow engines have improved since their introduction. With improvements in bearing technology the shaft speed of the engine was increased, greatly reducing the diameter of the centrifugal compressor. The short engine length remains an advantage of this design, particularly for use in helicopters where overall size is more important than frontal area. Also, its engine components are robust; axial-flow compressors are more liable to foreign object damage.

Although German designs were more advanced aerodynamically, the combination of simplicity and advanced British metallurgy meant that Whittle-derived designs were far more reliable than their German counterparts. British engines also were licensed widely in the US (see Tizard Mission),and were sold to the USSR who reverse engineered them with the Nene going on to power the famous MiG-15. American and Soviet designs, independent axial-flow types for the most part, would not come fully into their own until the 1960s, although the General Electric J47 provided excellent service in the F-86 Sabre in the 1950s.

By the 1950s the jet engine was almost universal in combat aircraft, with the exception of cargo, liaison and other specialty types. By this point some of the British designs were already cleared for civilian use, and had appeared on early models like the de Havilland Comet and Canadair Jetliner. By the 1960s all large civilian aircraft were also jet powered, leaving the piston engine in niche roles here as well.

Relentless improvements in the turboprop pushed the piston engine out of the mainstream entirely, leaving it serving only the smallest general aviation designs, and some use in drone aircraft. The ascension of the jet engine to almost universal use in aircraft took well under twenty years.

However, the story was not quite at an end, for the efficiency of turbojet engines was still rather worse than piston engines, but by the 1970s with the advent of high bypass jet engines, an innovation not foreseen by the early commentators like Edgar Buckingham, at high speeds and high altitudes that seemed absurd to them, only then did the fuel efficiency finally exceeded that of the best piston and propeller engines, and the dream of fast, safe, economical travel around the world finally arrived, and their dour, if well founded for the time, predictions that jet engines would never amount to much, killed forever.

Types

There are a large number of different types of jet engines, all of which achieve propulsion from a high speed exhaust jet.

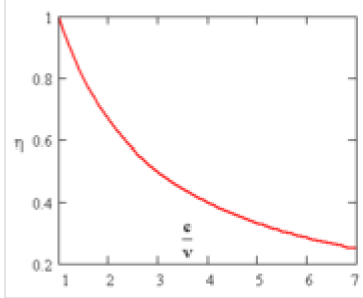
Type	Description	Advantages	Disadvantages
Water jet	For propelling boats; squirts water out the back through a nozzle	Can run in shallow water, high acceleration, no risk of engine overload (unlike propellers), less noise and vibration, highly manoeuvrable at all boat speeds, high speed efficiency, less vulnerable to damage from debris, very reliable, more load flexibility, less harmful to wildlife	Can be less efficient than a propeller at low speed, more expensive, higher weight in boat due to entrained water, will not perform well if boat is heavier than the jet is sized for
Motorjet	Most primitive airbreathing jet engine. Essentially a supercharged piston engine with a jet exhaust.	Higher exhaust velocity than a propeller, offering better thrust at high speed	Heavy, inefficient and underpowered
Turbojet	Generic term for simple turbine engine	Simplicity of design, efficient at supersonic speeds (~M2)	A basic design, misses many improvements in efficiency and power for subsonic flight, relatively noisy.
Low-bypass Turbofan	One- or two-stage fan added in front bypasses a proportion of the air through a bypass chamber surrounding the core. Compared with its turbojet ancestor, this allows for more efficient operation with somewhat less noise. This is the engine of high-speed military aircraft, some smaller private jets, and older civilian airliners such as the Boeing 707, the McDonnell Douglas DC-8, and their derivatives.	As with the turbojet, the design is aerodynamic, with only a modest increase in diameter over the turbojet required to accommodate the bypass fan and chamber. It is capable of supersonic speeds with minimal thrust drop-off at high speeds and altitudes yet still more efficient than the turbojet at subsonic operation.	Noisier and less efficient than high-bypass turbofan, with less static (Mach 0) thrust. Added complexity to accommodate dual shaft designs. More inefficient than a turbojet around M2 due to higher cross-sectional area.

High-bypass Turbofan	First stage compressor drastically enlarged to provide bypass airflow around engine core, and it provides significant amounts of thrust. Compared to the low-bypass turbofan and no-bypass turbojet, the high-bypass turbfan works on the principle of moving a great deal of air somewhat faster, rather than a small amount extremely fast. This translates into less noise. Most common form of jet engine in civilian use today- used in airliners like the Boeing 747, most 737s, and all Airbus aircraft.	Quieter due to greater mass flow and lower total exhaust speed, more efficient for a useful range of subsonic airspeeds for same reason, cooler exhaust temperature. High bypass variants exhibit good fuel economy.	Greater complexity (additional ducting, usually multiple shafts) and the need to contain heavy blades. Fan diameter can be extremely large, especially in high bypass turbofans such as the GE90. More subject to FOD and ice damage. Top speed is limited due to the potential for shockwaves to damage engine. Thrust lapse at higher speeds, which necessitates huge diameters and introduces additional drag.
Rocket	Carries all propellants and oxidants on-board, emits jet for propulsion	Very few moving parts, Mach 0 to Mach 25+, efficient at very high speed (> Mach 10.0 or so), thrust/weight ratio over 100, no complex air inlet, high compression ratio, very high speed (hypersonic) exhaust, good cost/thrust ratio, fairly easy to test, works in a vacuum-indeed works best exoatmospheric which is kinder on vehicle structure at high speed, fairly small surface area to keep cool, and no turbine in hot exhaust stream.	Needs lots of propellant- very low specific impulse — typically 100-450 seconds. Extreme thermal stresses of combustion chamber can make reuse harder. Typically requires carrying oxidiser on-board which increases risks. Extraordinarily noisy.
Ramjet	Intake air is compressed entirely by speed of oncoming air and duct shape (*divergent*)	Very few moving parts, Mach 0.8 to Mach 5+, efficient at high speed (> Mach 2.0 or so), lightest of all air-breathing jets (thrust/weight ratio up to 30 at optimum speed), cooling much easier than turbojets as no turbine blades to cool.	Must have a high initial speed to function, inefficient at slow speeds due to poor compression ratio, difficult to arrange shaft power for accessories, usually limited to a small range of speeds, intake flow must be slowed to subsonic speeds, noisy, fairly difficult to test, finicky to keep lit.
Turboprop (Turboshaft similar)	Strictly not a jet at all — a gas turbine engine is used as powerplant to drive propeller shaft (or rotor in the case of a helicopter)	High efficiency at lower subsonic airspeeds (300 knots plus), high shaft power to weight	Limited top speed (aeroplanes), somewhat noisy, complex transmission

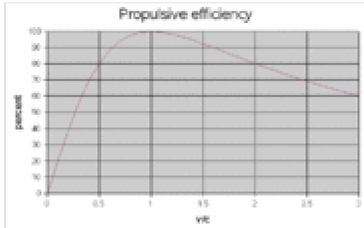
Propfan/Unducted Fan	Turboprop engine drives one or more propellers. Similar to a turbofan without the fan cowling.	Higher fuel efficiency, potentially less noisy than turbofans, could lead to higher-speed commercial aircraft, popular in the 1980s during fuel shortages	Development of propfan engines has been very limited, typically more noisy than turbofans, complexity
Pulsejet	Air is compressed and combusted intermittently instead of continuously. Some designs use valves.	Very simple design, commonly used on model aircraft	Noisy, inefficient (low compression ratio), works poorly on a large scale, valves on valved designs wear out quickly
Pulse detonation engine	Similar to a pulsejet, but combustion occurs as a detonation instead of a deflagration, may or may not need valves	Maximum theoretical engine efficiency	Extremely noisy, parts subject to extreme mechanical fatigue, hard to start detonation, not practical for current use
Air-augmented rocket	Essentially a ramjet where intake air is compressed and burnt with the exhaust from a rocket	Mach 0 to Mach 4.5+ (can also run exoatmospheric), good efficiency at Mach 2 to 4	Similar efficiency to rockets at low speed or exoatmospheric, inlet difficulties, a relatively undeveloped and unexplored type, cooling difficulties, very noisy, thrust/weight ratio is similar to ramjets.
Scramjet	Similar to a ramjet without a diffuser; airflow through the entire engine remains supersonic	Few mechanical parts, can operate at very high Mach numbers (Mach 8 to 15) with good efficiencies	Still in development stages, must have a very high initial speed to function (Mach >6), cooling difficulties, very poor thrust/weight ratio (~2), extreme aerodynamic complexity, airframe difficulties, testing difficulties/expense
Turborocket	A turbojet where an additional oxidizer such as oxygen is added to the airstream to increase maximum altitude	Very close to existing designs, operates in very high altitude, wide range of altitude and airspeed	Airspeed limited to same range as turbojet engine, carrying oxidizer like LOX can be dangerous. Much heavier than simple rockets.
Precooled jets / LACE	Intake air is chilled to very low temperatures at inlet in a heat exchanger before passing through a ramjet or turbojet engine. Can be combined with a rocket engine for orbital insertion.	Easily tested on ground. Very high thrust/weight ratios are possible (~14) together with good fuel efficiency over a wide range of airspeeds, mach 0-5.5+; this combination of efficiencies may permit launching to orbit, single stage, or very rapid, very long distance intercontinental	Exists only at the lab prototyping stage. Examples include RB545, SABRE, ATREX. Requires liquid hydrogen fuel which has very low density and heavily insulated tankage.

		travel.	

General physical principles



Dependence of the energy efficiency (η) from the exhaust speed/airplane speed ratio (c/v) for airbreathing jets



Dependence of the energy efficiency (η) upon the vehicle speed/exhaust speed ratio (v/c) for rocket engines

All jet engines are reaction engines that generate thrust by emitting a jet of fluid rearwards at relatively high speed. The forces on the inside of the engine needed to create this jet give a strong thrust on the engine which pushes the craft forwards.

Jet engines make their jet from propellant from tankage that is attached to the engine (as in a 'rocket') or from sucking in external fluid (very typically air) and expelling it at higher speed; or more commonly, a combination of the two sources.

Thrust

The motion impulse of the engine is equal to the fluid mass multiplied by the speed at which the engine emits this mass:

$$I = m\,c$$

where m is the fluid mass per second and c is the exhaust speed. In other words, a vehicle gets the same thrust if it outputs a lot of exhaust very slowly, or a little exhaust very quickly.

However, when an vehicle moves with certain velocity v, the fluid moves towards it, creating an opposing ram drag at the intake:

$$m\,v$$

Most types of jet engine have an intake, which provides the bulk of the fluid exiting the exhaust. Conventional rocket motors, however, do not have an intake, the oxidizer and fuel both being carried within the vehicle. Therefore, rocket motors do not have ram drag; the gross thrust of the nozzle is the net thrust of the engine. Consequently, the thrust characteristics of a rocket motor are completely different from that of an air breathing jet engine.

The jet engine with an intake is only useful if the velocity of the gas from the engine, c, is greater than the vehicle velocity, v, as the net engine thrust is the same as if the gas were emitted with the velocity c-v. So the thrust is actually equal to

$$S = m\,(c-v)$$

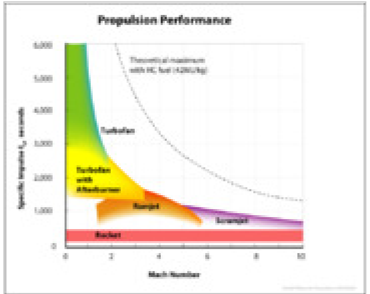
Energy efficiency

For all jet engines the *propulsive efficiency* (essentially energy efficiency) is highest when the engine emits an exhaust jet at a speed that is the same as, or nearly the same as, the vehicle velocity. The exact formula for air-breathing engines as given in the literature, is

$$\eta_p = \frac{2}{1 + \frac{c}{v}}$$

A corollary of this is that, particularly in air breathing engines, it is more energy efficient to accelerate a large amount of air by a little bit than a small amount by a large amount, even though the thrust is the same.

In addition to propulsive efficiency, another factor is cycle efficiency; essentially a jet engine is typically a form of heat engine. Heat engine efficiency is determined by the ratio of temperatures that are reached in the engine to that they are exhausted at from the nozzle, which in turn is limited by the overall pressure ratio that can be achieved.



Specific impulse as a function of speed for different jet types with kerosene fuel (hydrogen I_{sp} would be about twice as high). Although efficiency plummets with speed, greater distances are covered, it turns out that efficiency per unit distance (per km or mile) is roughly independent of speed for jet engines as a group; however airframes become inefficient at supersonic speeds

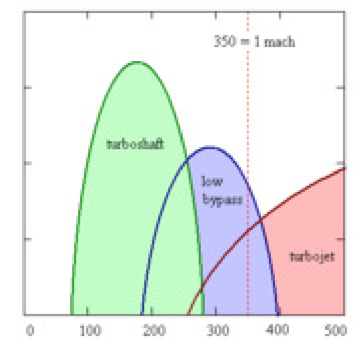
Fuel/propellant consumption

A closely related (but different) concept to energy efficiency is propellant consumption. Propellant consumption in jet engines is measured by **Specific Fuel Consumption**, **Specific impulse** or **Effective exhaust velocity**. They all measure the same thing, specific impulse and effective exhaust velocity are strictly proportional, whereas specific fuel consumption is inversely proportional to the others.

For airbreathing engines such as turbojets energy efficiency and propellant (fuel) efficiency are much the same thing, since the propellant is the source of energy- a fuel. In rocketry, the propellant is also the exhaust, and this means that a high energy propellant gives better propellant efficiency but *lower* energy efficiency.

Comparison of types

Turboprops obtain little thrust from jet effect, but are useful for comparison. They are gas turbine engines that have a rotating fan that takes and accelerates the large mass of air but by a relatively small change in speed. This low speed limits the speed of any propeller driven airplane. When the plane speed exceeds this limit, propellers no longer provide any thrust (c-v < 0). However, because they accelerate a large mass of air, turboprops are very efficient.



Comparative suitability for (left to right) turboshaft, low bypass and turbojet to fly at 10 km altitude in various speeds. Horizontal axis - speed, m/s. Vertical axis displays engine efficiency.

turbojets and other similar engines accelerate a much smaller mass of the air and burned fuel, but they emit it at the much higher speeds possible with a de Laval nozzle. This is why they are suitable for supersonic and higher speeds.

Low bypass turbofans have the mixed exhaust of the two air flows, running at different speeds (c1 and c2). The thrust of such engine is

$$S = m1\ (c1 - v) + m2\ (c2 - v)$$

where m1 and m2 are the air masses, being blown from the both exhausts. Such engines are effective at lower speeds, than the pure jets, but at higher speeds than the turboshafts and propellers in general. For instance, at the 10 km altitude, turboshafts are most effective at about 0.4 mach, low bypass turbofans become more effective at about 0.75 mach and turbojets become more effective as mixed exhaust engines when the speed approaches 2-3 mach - 2-3x the speed of sound.

Rocket engines have extremely high exhaust velocity and thus are best suited for high speeds (hypersonic) and great altitudes. At any given throttle, the thrust and efficiency of a rocket motor improves slightly with increasing altitude (because the back-pressure falls thus increasing net thrust at the nozzle exit plane), whereas with a turbojet (or turbofan) the falling density of the air entering the intake (and the hot gases leaving the nozzle) causes the net thrust to decrease with increasing altitude. Rocket engines are more efficient than even scramjets above roughly Mach 15.

Noise

Noise is due to shockwaves that form when the exhaust jet interacts with the external air.

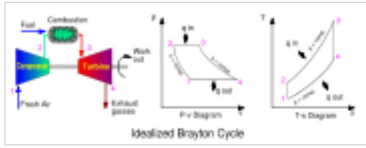
The intensity of the noise is proportional to the thrust as well as proportional to the fourth power of the jet velocity.

Generally then, the lower speed exhaust jets emitted from engines such as high bypass turbofans are the quietest, whereas the fastest jets are the loudest.

Although some variation in jet speed can often be arranged from a jet engine (such as by throttling back and adjusting the nozzle) it is difficult to vary the jet speed from an engine over a very wide range. Therefore since engines for supersonic vehicles such as Concorde, military jets and rockets inherently need to have supersonic exhaust at top speed, so these vehicles are especially noisy even at low speeds.

Common types

Turbojet engines


A turbojet engine, in its simplest form is simply a gas turbine with a nozzle attached

A turbojet engine is a type of internal combustion engine often used to propel aircraft. Air is drawn into the rotating compressor via the intake and is compressed, through successive stages, to a higher pressure before entering the combustion chamber. Fuel is mixed with the compressed air and ignited by flame in the eddy of a flame holder. This combustion process significantly raises the temperature and volume of the air. Hot combustion products leaving the combustor expand through a gas turbine, where power is extracted to drive the compressor. This expansion process reduces both the gas temperature and pressure but sufficient fuel is burnt so that both parameters are usually still well above ambient conditions at exit from the turbine. The gas stream is then expanded to ambient pressure via a propelling nozzle, producing a high velocity jet as the exhaust. If the jet velocity exceeds the aircraft flight velocity, there is a net forward thrust upon the airframe.

Under normal circumstances, the pumping action of the compressor prevents any backflow, thus facilitating the continuous-flow process of the engine. Indeed, the entire process is similar to a four-stroke cycle, but with induction, compression, ignition, expansion and exhaust taking place simultaneously, but in different sections of the engine. The efficiency of a jet engine is strongly dependent upon the overall pressure ratio (combustor entry pressure/intake delivery pressure) and the turbine inlet temperature of the cycle.

It is also perhaps instructive to compare turbojet engines with propeller engines. Turbojet engines take a relatively small mass of air and accelerate it by a large amount, whereas a propeller takes a large mass of air and accelerates it by a small amount. The high-speed exhaust of a turbojet engine makes it efficient at high speeds (especially supersonic speeds) and high altitudes; Concorde used this type for example. On slower aircraft and those required to fly short stages, a gas turbine-powered propeller engine, commonly known as a turboprop, is more common and much more efficient. Very small aircraft generally use conventional piston engines to drive a propeller but small turboprops are getting smaller as engineering technology improves.

The turbojet described above is a single-spool design, in which a single shaft connects the turbine to the compressor. Two spool designs have two concentric turbine-compressor systems, that spin independently with the turbine and compressors for each section connected from opposite ends of the engine via concentric shafts. This allows for a higher compression ratio as well as improved compressor stability during engine throttle movements. Three spool designs also exist.

Turbofan engines

Most modern jet engines are actually turbofans, where the low pressure compressor acts as a fan, supplying supercharged air not only to the engine core, but to a bypass duct. The bypass airflow either passes to a separate 'cold nozzle' or mixes with low pressure turbine exhaust gases, before expanding through a 'mixed flow nozzle'.

Turbofans are used for airliners because they give an exhaust speed that is better matched for subsonic airliners, at airliners flight speed conventional turbojet engines generate an exhaust that ends up travelling very fast backwards, and this wastes energy. By emitting the exhaust so that it ends up travelling more slowly, better fuel consumption is achieved. In addition, the lower exhaust speed gives much lower noise.

In the 1960s there was little difference between civil and military jet engines, apart from the use of afterburning in some (supersonic) applications. Civil turbofans today have a low exhaust speed (low *specific thrust* -net thrust divided by airflow) to keep jet noise to a minimum and to improve fuel efficiency. Consequently the bypass ratio (bypass flow divided by core flow) is relatively high (ratios from 4:1 up to 8:1 are common). Only a single fan stage is required, because a low specific thrust implies a low fan pressure ratio.

Today's military turbofans, however, have a relatively high specific thrust, to maximize the thrust for a given frontal area, jet noise being of less concern in military uses relative to civil uses. Multistage fans are normally needed to reach the relatively high fan pressure ratio needed for high specific thrust. Although high turbine inlet temperatures are often employed, the bypass ratio tends to be low, usually significantly less than 2.0.

An approximate equation for calculating the net thrust of a jet engine, be it a turbojet or a mixed turbofan, is:

$$F_n = \dot{m}(V_{jfe} - V_a)$$

where:

$\dot{m}$ = intake mass flow rate

V_{jfe} = fully expanded jet velocity (in the exhaust plume)

V_a = aircraft flight velocity

While the $\dot{m}.V_{jfe}$ term represents the gross thrust of the nozzle, the $\dot{m}.V_a$ term represents the ram drag of the intake.

Rocket engines

The third most common form of jet engine is the rocket engine.

Rocket engines are used for rockets because their extremely high exhaust velocity and independence from the atmospheric oxygen permits them to achieve spaceflight.

This is used for launching satellites, space exploration and manned access, and permitted landing on the moon in 1969.

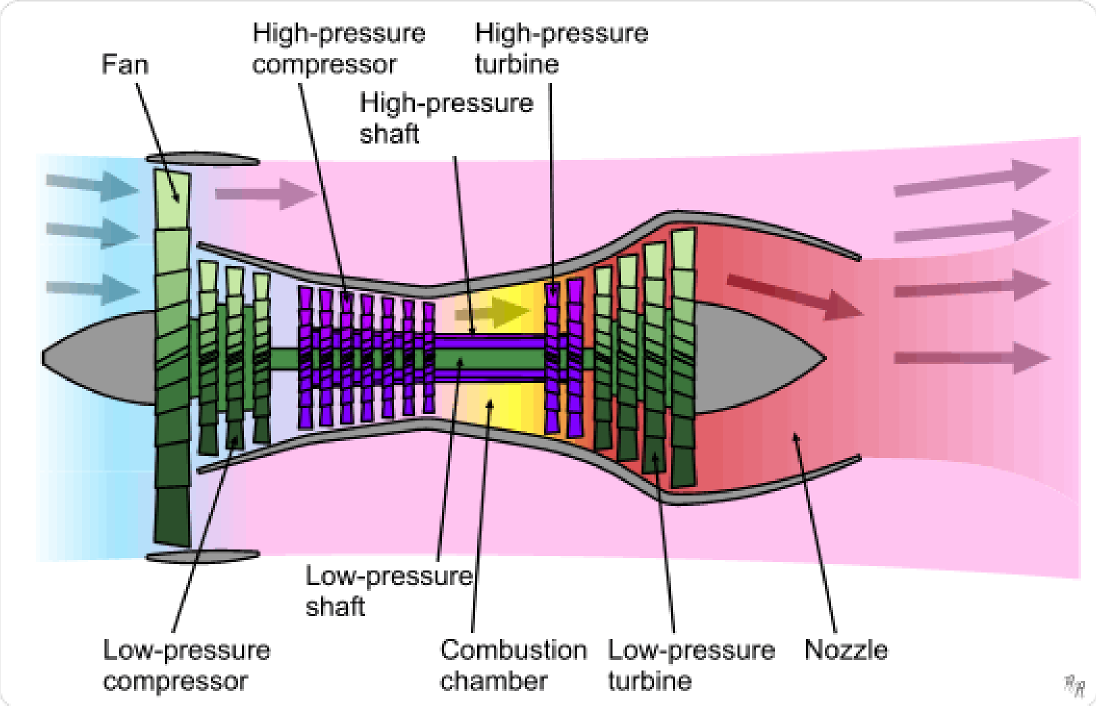
However, the high exhaust speed and the heavier propellant mass results in less efficient flight than turbojets, and their use is largely restricted to very high altitudes or where very high accelerations are needed as rocket engines themselves have a very high thrust-to-weight ratio.

An approximate equation for the net thrust of a rocket engine is:

$$F = \dot{m} g_0 I_{sp-vac} - A_e P$$

Where F is the thrust, $I_{sp(vac)}$ is the specific impulse, g_0 is a standard gravity, Ae is the area of the exhaust bell at the exit and P is the atmospheric pressure.

Major components



The major components of a jet engine are similar across the major different types of engines, although not all engine types have all components. The major parts include:

- **Cold Section:**
  - **Air intake (Inlet)** — The standard reference frame for a jet engine is the aircraft itself. For subsonic aircraft, the air intake to a jet engine presents no special difficulties, and consists essentially of an opening which is designed to minimise drag, as with any other aircraft component. However, the air reaching the compressor of a normal jet engine must be travelling below the speed of sound, even for supersonic aircraft, to sustain the flow mechanics of the compressor and turbine blades. At supersonic flight speeds, shockwaves form in the intake system and reduce the recovered pressure at inlet to the compressor. So some supersonic intakes use devices, such as a cone or ramp, to increase pressure recovery, by making more efficient use of the shock wave system.

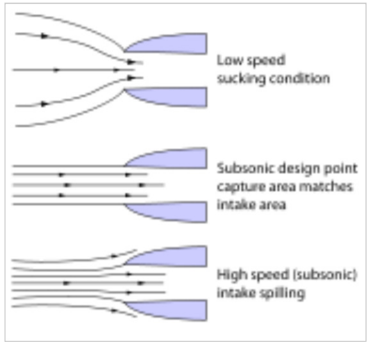
  - **Compressor** or **Fan** — The compressor is made up of stages. Each stage consists of vanes which rotate, and stators which remain stationary. As air is drawn deeper through the compressor, its heat and pressure increases. Energy is derived from the **turbine** (see below), passed along the **shaft**.
  - **Bypass ducts** much of the thrust of essentially all modern jet engines comes from air from the front compressor that bypasses the combustion chamber and gas turbine section that leads directly to the nozzle or afterburner (where fitted).
- **Common:**
  - **Shaft** — The shaft connects the **turbine** to the **compressor**, and runs most of the length of the engine. There may be as many as three concentric shafts, rotating at independent speeds, with as many sets of turbines and compressors. Other services, like a bleed of cool air, may also run down the shaft.
- **Hot section:**
  - **Combustor** or **Can** or **Flameholders** or **Combustion Chamber** — This is a chamber where fuel is continuously burned in the compressed air.
  - **Turbine** — The turbine is a series of bladed discs that act like a windmill, gaining energy from the hot gases leaving the **combustor**. Some of this energy is used to drive the **compressor**, and in some turbine engines (ie turboprop, turboshaft or turbofan engines), energy is extracted by additional turbine discs and used to drive devices such as propellers, bypass fans or helicopter rotors. One type, a **free turbine**, is configured such that the turbine disc driving the compressor rotates independently of the discs that power the external components. Relatively cool air, bled from the compressor, may be used to cool the turbine blades and vanes, to prevent them from melting.
  - **Afterburner** or **reheat** (chiefly UK) — (mainly military) Produces extra thrust by burning extra fuel, usually inefficiently, to significantly raise Nozzle Entry Temperature at the **exhaust**. Owing to a larger volume flow (i.e. lower density) at exit from the afterburner, an increased nozzle flow area is required, to maintain satisfactory engine matching, when the afterburner is alight.
  - **Exhaust** or **Nozzle** — Hot gases leaving the engine exhaust to atmospheric pressure via a nozzle, the objective being to produce a high velocity jet. In most cases, the nozzle is convergent and of fixed flow area.
  - **Supersonic nozzle** — If the Nozzle Pressure Ratio (Nozzle Entry Pressure/Ambient Pressure) is very high, to maximize thrust it may be worthwhile, despite the additional weight, to fit a convergent-divergent (de Laval) nozzle. As the name suggests, initially this type of nozzle is convergent, but beyond the throat (smallest flow area), the flow area starts to increase to form the divergent portion. The expansion to atmospheric pressure and supersonic gas velocity continues downstream of the throat, whereas in a convergent nozzle the expansion beyond sonic velocity occurs externally, in the exhaust plume. The former process is more efficient than the latter.

The various components named above have constraints on how they are put together to generate the most efficiency or performance. The performance and efficiency of an engine can never be taken in isolation; for example fuel/distance efficiency of a supersonic jet engine maximises at about mach 2, whereas the drag for the vehicle carrying it is increasing as a square law and has much extra drag in the transonic region. The highest fuel efficiency for the overall vehicle is thus typically at Mach ~0.85.

For the engine optimisation for its intended use, important here is air intake design, overall size, number of compressor stages (sets of blades), fuel type, number of exhaust stages, metallurgy of components, amount of bypass air used, where the bypass air is introduced, and many other factors. For instance, let us consider design of the air intake.

Air intakes

Subsonic inlets



Pitot intake operating modes

Pitot intakes are the dominant type for subsonic applications. A subsonic pitot inlet is little more than a tube with an aerodynamic fairing around it.

At zero airspeed (i.e., rest), air approaches the intake from a multitude of directions: from directly ahead, radially, or even from behind the plane of the intake lip.

At low airspeeds, the streamtube approaching the lip is larger in cross-section than the lip flow area, whereas at the intake design flight Mach number the two flow areas are equal. At high flight speeds the streamtube is smaller, with excess air spilling over the lip.

Beginning around 0.85 Mach, shock waves can occur as the air accelerates through the intake throat.

Careful radiusing of the lip region is required to optimize intake pressure recovery (and distortion) throughout the flight envelope.

Supersonic inlets

Supersonic intakes exploit shock waves to decelerate the airflow to a subsonic condition at compressor entry.

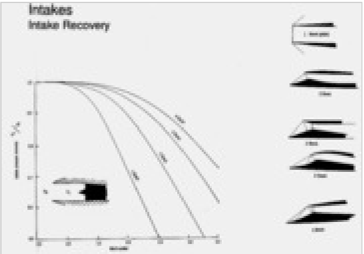
There are basically two forms of shock waves:

1) Normal shock waves lie perpendicular to the direction of the flow. These form sharp fronts and shock the flow to subsonic speeds. Microscopically the air molecules smash into the subsonic crowd of molecules like alpha rays. Normal shock waves tend to cause a large drop in stagnation pressure. Basically, the higher the supersonic entry Mach number to a normal shock wave, the lower the subsonic exit Mach number and the stronger the shock (i.e. the greater the loss in stagnation pressure across the shock wave).

2) Conical (3-dimensional) and oblique shock waves (2D) are angled rearwards, like the bow wave on a ship or boat, and radiate from a flow disturbance such as a cone or a ramp. For a given inlet Mach number, they are weaker than the equivalent normal shock wave and, although the flow slows down, it remains supersonic throughout. Conical and oblique shock waves turn the flow, which continues in the new direction, until another flow disturbance is encountered downstream.

Note: Comments made regarding 3 dimensional conical shock waves, generally also apply to 2D oblique shock waves.

A sharp-lipped version of the pitot intake, described above for subsonic applications, performs quite well at moderate supersonic flight speeds. A detached normal shock wave forms just ahead of the intake lip and 'shocks' the flow down to a subsonic velocity. However, as flight speed increases, the shock wave becomes stronger, causing a larger percentage decrease in stagnation pressure (i.e. poorer pressure recovery). An early US supersonic fighter, the F-100 Super Sabre, used such an intake.



An unswept lip generate a shock wave, which is reflected multiple times in the inlet. The more reflections before the flow gets subsonic, the better pressure recovery

More advanced supersonic intakes, excluding pitots:

a) exploit a combination of conical shock wave/s and a normal shock wave to improve pressure recovery at high supersonic flight speeds. Conical shock wave/s are used to reduce the supersonic Mach number at entry to the normal shock wave, thereby reducing the resultant overall shock losses.

b) have a design shock-on-lip flight Mach number, where the conical/oblique shock wave/s intercept the cowl lip, thus enabling the streamtube capture area to equal the intake lip area. However, below the shock-on-lip flight Mach number, the shock wave angle/s are less oblique, causing the streamline approaching the lip to be deflected by the presence of the cone/ramp. Consequently, the intake capture area is less than the intake lip area, which reduces the intake airflow. Depending on the airflow characteristics of the engine, it may be desirable to lower the ramp angle or move the cone rearwards to refocus the shockwaves onto the cowl lip to maximise intake airflow.

c) are designed to have a normal shock in the ducting downstream of intake lip, so that the flow at compressor/fan entry is always subsonic. However, if the engine is throttled back, there is a reduction in the corrected airflow of the LP compressor/fan, but (at supersonic conditions) the corrected airflow at the intake lip remains constant, because it is determined by the flight Mach number and intake incidence/yaw. This discontinuity is overcome by the normal shock moving to a lower cross-sectional area in the ducting, to decrease the Mach number at entry to the shockwave. This weakens the shockwave, improving the overall intake pressure recovery. So, the absolute airflow stays constant, whilst the corrected airflow at compressor entry falls (because of a higher entry pressure). Excess intake airflow may also be dumped overboard or into the exhaust system, to prevent the conical/oblique shock waves being disturbed by the normal shock being forced too far forward by engine throttling.

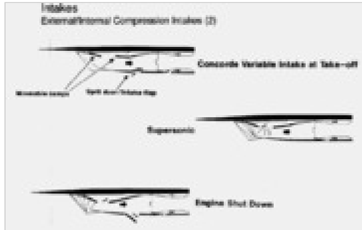
Many second generation supersonic fighter aircraft featured an inlet cone, which was used to form the conical shock wave. This type of inlet cone is clearly seen at the very front of the English Electric Lightning and MiG-21 aircraft, for example.

The same approach can be used for air intakes mounted at the side of the fuselage, where a half cone serves the same purpose with a semicircular air intake, as seen on the F-104 Starfighter and BAC TSR-2.

Some intakes are biconic; that is they feature two conical surfaces: the first cone is supplemented by a second, less oblique, conical surface, which generates an extra conical shockwave, radiating from the junction between the two cones. A biconic intake is usually more efficient than the equivalent conical intake, because the entry Mach number to the normal shock is reduced by the presence of the second conical shock wave.

A very sophisticated conical intake was featured on the SR-71's Pratt & Whitney J58s that could move a conical spike fore and aft within the engine nacelle, preventing the shockwave formed on the spike from entering the engine and stalling the engine, while keeping it close enough to give good compression. Movable cones are uncommon.

A more sophisticated design than cones is to angle the intake so that one of its edges forms a ramp. An oblique shockwave will form at the start of the ramp. The Century Series of US jets featured several variants of this approach, usually with the ramp at the outer vertical edge of the intake, which was then angled back inward towards the fuselage. Typical examples include the Republic F-105 Thunderchief and F-4 Phantom.

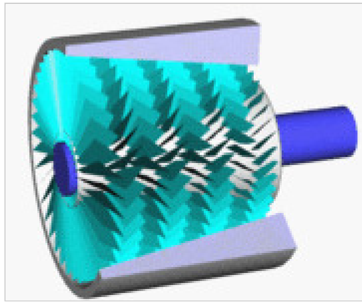


Concorde intake operating modes

Later this evolved so that the ramp was at the top horizontal edge rather than the outer vertical edge, with a pronounced angle downwards and rearwards. This design simplified the construction of intakes and allowed use of variable ramps to control airflow into the engine. Most designs since the early 1960s now feature this style of intake, for example the F-14 Tomcat, Panavia Tornado and Concorde.

From another point of view, like in a supersonic nozzle the corrected (or non-dimensional) flow has to be the same at the intake lip, at the intake throat and at the turbine. One of this three can be fixed. For inlets the throat is made variable and some air is bypassed around the turbine and directly fed into the afterburner. Unlike in a nozzle the inlet is either unstable or inefficient, because a normal shock wave in the throat will suddenly move to the lip, thereby increasing the pressure at the lip, leading to drag and reducing the pressure recovery, leading to turbine surge and the loss of one SR-71.

Compressors


Axial compressors

Axial compressors rely on spinning blades that have aerofoil sections, similar to aeroplane wings. As with aeroplane wings in some conditions the blades can stall. If this happens, the airflow around the stalled compressor can reverse direction violently. Each design of a compressor has an associated operating map of airflow versus rotational speed for characteristics peculiar to that type (see compressor map).

At a given throttle condition, the compressor operates somewhere along the steady state running line. Unfortunately, this operating line is displaced during transients. Many compressors are fitted with anti-stall systems in the form of bleed bands or variable geometry stators to decrease the likelihood of surge. Another method is to split the compressor into two or more units, operating on separate concentric shafts.


Compressor stage GE J79

Another design consideration is the average stage loading. This can be kept at a sensible level either by increasing the number of compression stages (more weight/cost) or the mean blade speed (more blade/disc stress).

Although large flow compressors are usually all-axial, the rear stages on smaller units are too small to be robust. Consequently, these stages are often replaced by a single centrifugal unit. Very small flow compressors often employ two centrifugal compressors, connected in series. Although in isolation centrifugal compressors are capable of running at quite high pressure ratios (e.g. 10:1), impeller stress considerations limit the pressure ratio that can be employed in high overall pressure ratio engine cycles.

Increasing overall pressure ratio implies raising the high pressure compressor exit temperature. This implies a higher high pressure shaft speed, to maintain the datum blade tip Mach number on the rear compressor stage. Stress considerations, however, may limit the shaft speed increase, causing the original compressor to throttle-back aerodynamically to a lower pressure ratio than datum.

Combustors


Combustion chamber GE J79

Great care must be taken to keep the flame burning in a moderately fast moving airstream, at all throttle conditions, as efficiently as possible. Since the turbine cannot withstand stoichiometric temperatures (a mixture ratio of around 15:1), some of the compressor air is used to quench the exit temperature of the combustor to an acceptable level (an overall mixture ratio of between 45:1 and 130:1 is used). Air used for combustion is considered to be primary airflow, while excess air used for cooling is called secondary airflow. Combustor configurations include can, annular, and can-annular.

Turbines


Turbine Stage GE J79

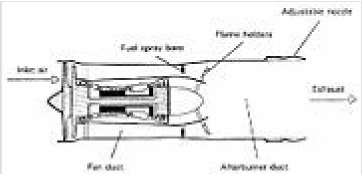
Because a turbine expands from high to low pressure, there is no such thing as turbine surge or stall. The turbine needs fewer stages than the compressor, mainly because the higher inlet temperature reduces the deltaT/T (and thereby the pressure ratio) of the expansion process. The blades have more curvature and the gas stream velocities are higher.

Designers must, however, prevent the turbine blades and vanes from melting in a very high temperature and stress environment. Consequently bleed air extracted from the compression system is often used to cool the turbine blades/vanes internally. Other solutions are improved materials and/or special insulating coatings. The discs must be specially shaped to withstand the huge stresses imposed by the rotating blades. They take the form of impulse, reaction, or combination impulse-reaction shapes. Improved materials help to keep disc weight down.

Turbopumps

Turbopumps are centrifugal pumps which are spun by gas turbines and are used to raise the propellant pressure above the pressure in the combustion chamber so that it can be injected and burnt. Turbopumps are very commonly used with rockets, but ramjets and turbojets also have been known to use them.

Afterburners (reheat)



Turbofan fitted with afterburner

Due to temperature limitations with the gas turbines, jet engines do not consume all the oxygen in the air ('run stoichiometric'). Afterburners burn the remaining oxygen after exiting the turbines, but usually do so inefficiently due to the low pressures typically found at this part of the jet engine; however this gains significant thrust, which can be useful. Engines intended for extended use with afterburners often have variable nozzles and other details.

Nozzles

Afterburner GE J79

The primary objective of a nozzle is to expand the exhaust stream to atmospheric pressure, and form it into a high speed jet to propel the vehicle. For airbreathing engines, if the fully expanded jet has a higher speed than the aircraft's airspeed, then there is a net rearward momentum gain to the air and there will be a forward thrust on the airframe.

Simple convergent nozzles are used on many jet engines. If the nozzle pressure ratio is above the critical value (about 1.8:1) a convergent nozzle will choke, resulting in some of the expansion to atmospheric pressure taking place downstream of the throat (i.e. smallest flow area), in the jet wake. Although much of the gross thrust produced will still be from the jet momentum, additional (pressure) thrust will come from the imbalance between the throat static pressure and atmospheric pressure.

Many military combat engines incorporate an afterburner (or reheat) in the engine exhaust system. When the system is lit, the nozzle throat area must be increased, to accommodate the extra exhaust volume flow, so that the turbomachinery is unaware that the afterburner is lit. A variable throat area is achieved by moving a series of overlapping petals, which approximate the circular nozzle cross-section.

At high nozzle pressure ratios, the exit pressure is often above ambient and much of the expansion will take place downstream of a convergent nozzle, which is inefficient. Consequently, some jet engines (notably rockets) incorporate a convergent-divergent nozzle, to allow most of the expansion to take place against the inside of a nozzle to maximise thrust. However, unlike the fixed con-di nozzle used on a conventional rocket motor, when such a device is used on a turbojet engine it has to be a complex variable geometry device, to cope with the wide variation in nozzle pressure ratio encountered in flight and engine throttling. This further increases the weight and cost of such an installation.


Variable Exhaust Nozzle, on the GE F404-400 low-bypass turbofan installed on a Boeing F/A-18 Hornet

The simpler of the two is the **ejector nozzle**, which creates an effective nozzle through a secondary airflow and spring-loaded petals. At subsonic speeds, the airflow constricts the exhaust to a convergent shape. As the aircraft speeds up, the two nozzles dilate, which allows the exhaust to form a convergent-divergent shape, speeding the exhaust gasses past Mach 1. More complex engines can actually use a tertiary airflow to reduce exit area at very low speeds. Advantages of the ejector nozzle are relative simplicity and reliability. Disadvantages are average performance (compared to the other nozzle type) and relatively high drag due to the secondary airflow. Notable aircraft to have utilized this type of nozzle include the SR-71, Concorde, F-111, and Saab Viggen

For higher performance, it is necessary to use an **iris nozzle**. This type uses overlapping, hydraulically adjustable "petals". Although more complex than the ejector nozzle, it has significantly higher performance and smoother airflow. As such, it is employed primarily on high-performance fighters such as the F-14, F-15, F-16, though is also used in high-speed bombers such as the B-1B. Some modern iris nozzles additionally have the ability to change the angle of the thrust (see thrust vectoring).


Iris vectored thrust nozzle

Rocket motors also employ convergent-divergent nozzles, but these are usually of fixed geometry, to minimize weight. Because of the much higher nozzle pressure ratios experienced, rocket motor con-di nozzles have a much greater area ratio (exit/throat) than those fitted to jet engines. The Convair F-106 Delta Dart has used such a nozzle design, as part of its overall design specification as a aerospace interceptor for high-altitude bomber interception, where conventional nozzle design would prove ineffective.

At the other extreme, some high bypass ratio civil turbofans use an extremely low area ratio (less than 1.01 area ratio), convergent-divergent, nozzle on the bypass (or mixed exhaust) stream, to control the fan working line. The nozzle acts as if it has variable geometry. At low flight speeds the nozzle is unchoked (less than a Mach number of unity), so the exhaust gas speeds up as it approaches the throat and then slows down slightly as it reaches the divergent section. Consequently, the nozzle exit area controls the fan match and, being larger than the throat, pulls the fan working line slightly away from surge. At higher flight speeds, the ram rise in the intake increases nozzle pressure ratio to the point where the throat becomes choked (M=1.0). Under these circumstances, the throat area dictates the fan match and being smaller than the exit pushes the fan working line slightly towards surge. This is not a problem, since fan surge margin is much better at high flight speeds.

Thrust reversers

These either consist of cups that swing across the end of the nozzle and deflect the jet thrust forwards (as in the DC-9), or they are two panels behind the cowling that slide backward and reverse only the fan thrust (the fan produces the majority of the thrust). This is the case on many large aircraft such as the 747, C-17, KC-135, etc.

Cooling systems

All jet engines require high temperature gas for good efficiency, typically achieved by combusting hydrocarbon or hydrogen fuel. Combustion temperatures can be as high as 3500K (5841F) in rockets, far above the melting point of most materials, but normal airbreathing jet engines use rather lower temperatures.

Cooling systems are employed to keep the temperature of the solid parts below the failure temperature.

Air systems

A complex around combustor and is injected into the rim of the rotating turbine disc. The cooling air then passes through complex passages within the turbine blades. After removing heat from the blade material, the air (now fairly hot) is vented, via cooling holes, into the main gas stream. Cooling air for the turbine vanes undergoes a similar process.

Cooling the leading edge of the blade can be difficult, because the pressure of the cooling air just inside the cooling hole may not be much different from that of the oncoming gas stream. One solution is to incorporate a cover plate on the disc. This acts as a centrifugal compressor to pressurize the cooling air before it enters the blade. Another solution is to use an ultra-efficient turbine rim seal to pressurize the area where the cooling air passes across to the rotating disc.

Seals are used to prevent oil leakage, control air for cooling and prevent stray air flows into turbine cavities.

A series of (e.g. labyrinth) seals allow a small flow of bleed air to wash the turbine disc to extract heat and, at the same time, pressurize the turbine rim seal, to prevent hot gases entering the inner part of the engine. Other types of seals are hydraulic, brush, carbon etc.

Small quantities of compressor bleed air are also used to cool the shaft, turbine shrouds, etc. Some air is also used to keep the temperature of the combustion chamber walls below critical. This is done using primary and secondary airholes which allow a thin layer of air to cover the inner walls of the chamber preventing excessive heating.

Exit temperature is dependent on the turbine upper temperature limit depending on the material. Reducing the temperature will also prevent thermal fatigue and hence failure. Accessories may also need their own cooling systems using air from the compressor or outside air.

Air from compressor stages is also used for heating of the fan, airframe anti-icing and for cabin heat. Which stage is bled from depends on the atmospheric conditions at that altitude.

Fuel system

Apart from providing fuel to the engine, the fuel system is also used to control propeller speeds, compressor airflow and cool lubrication oil. Fuel is usually introduced by an atomized spray, the amount of which is controlled automatically depending on the rate of airflow.

So the sequence of events for increasing thrust is, the throttle opens and fuel spray pressure is increased, increasing the amount of fuel being burned. This means that exhaust gases are hotter and so are ejected at higher acceleration, which means they exert higher forces and therefore increase the engine thrust directly. It also increases the energy extracted by the turbine which drives the compressor even faster and so there is an increase in air flowing into the engine as well.

Obviously, it is the rate of the **mass** of the airflow that matters since it is the change in momentum (mass x velocity) that produces the force. However, density varies with altitude and hence inflow of mass will also vary with altitude, temperature etc. which means that throttle values will vary according to all these parameters without changing them manually.

This is why fuel flow is controlled automatically. Usually there are 2 systems, one to control the pressure and the other to control the flow. The inputs are usually from pressure and temperature probes from the intake and at various points through the engine. Also throttle inputs, engine speed etc. are required. These affect the high pressure fuel pump.

Fuel control unit (FCU)

This element is something like a mechanical computer. It determines the output of the fuel pump by a system of valves which can change the pressure used to cause the pump stroke, thereby varying the amount of flow.

Take the possibility of increased altitude where there will be reduced air intake pressure. In this case, the chamber within the FCU will expand which causes the spill valve to bleed more fuel. This causes the pump to deliver less fuel until the opposing chamber pressure is equivalent to the air pressure and the spill valve goes back to its position.

When the throttle is opened, it releases i.e. lessens the pressure which lets the throttle valve fall. The pressure is transmitted (because of a back-pressure valve i.e. no air gaps in fuel flow) which closes the FCU spill valves (as they are commonly called) which then increases the pressure and causes a higher flow rate.

The engine speed governor is used to prevent the engine from over-speeding. It has the capability of disregarding the FCU control. It does this by use of a diaphragm which senses the engine speed in terms of the centrifugal pressure caused by the rotating rotor of the pump. At a critical value, this diaphragm causes another spill valve to open and bleed away the fuel flow.

There are other ways of controlling fuel flow for example with the dash-pot throttle lever. The throttle has a gear which meshes with the control valve (like a rack and pinion) causing it to slide along a cylinder which has ports at various positions. Moving the throttle and hence sliding the valve along the cylinder, opens and closes these ports as designed. There are actually 2 valves viz. the throttle and the control valve. The control valve is used to control pressure on one side of the throttle valve such that it gives the right opposition to the throttle control pressure. It does this by controlling the fuel outlet from within the cylinder.

So for example, if the throttle valve is moved up to let more fuel in, it will mean that the throttle valve has moved into a position which allows more fuel to flow through and on the other side, the required pressure ports are opened to keep the pressure balance so that the throttle lever stays where it is.

At initial acceleration, more fuel is required and the unit is adapted to allow more fuel to flow by opening other ports at a particular throttle position. Changes in pressure of outside air i.e. altitude, speed of aircraft etc are sensed by an air capsule.

Fuel pump

Fuel pumps are used to raise the fuel pressure above the pressure in the combustion chamber so that the fuel can be injected. Fuel pumps are usually driven by the main shaft, via gearing.

Turbopumps are very commonly used with liquid-fuelled rockets and rely on the expansion of an onboard gas through a turbine.

Ramjet turbopumps use ram air expanding through a turbine.

Engine starting system

The fuel system as explained above, is one of the 2 systems required for starting the engine. The other is the actual ignition of the air/fuel mixture in the chamber. Usually, an auxiliary power unit is used to start the engines. It has a starter motor which has a high torque transmitted to the compressor unit. When the optimum speed is reached, i.e. the flow of gas through the turbine is sufficient, the turbines take over. There are a number of different starting methods such as *electric*, *hydraulic*, *pneumatic* etc.

The **electric** starter works with gears and clutch plate linking the motor and the engine. The clutch is used to disengage when optimum speed is achieved. This is usually done automatically. The electric supply is used to start the motor as well as for ignition. The voltage is usually built up slowly as starter gains speed.

Some military aircraft need to be started quicker than the electric method permits and hence they use other methods such as a turbine starter. This is an impulse turbine impacted by burning gases from a cartridge. It is geared to rotate the engine and also connected to an automatic disconnect system. The cartridge is set alight electrically and used to turn the turbine.

Another turbine starter system is almost exactly like a little engine. Again the turbine is connected to the engine via gears. However, the turbine is turned by burning gases - usually the fuel is isopropyl nitrate stored in a tank and sprayed into a combustion chamber. Again, it is ignited with a spark plug. Everything is electrically controlled, such as speed etc.

Most Commercial aircraft and large Military Transport airplanes usually use what is called an **auxiliary power unit** or **APU**. It is normally a small gas turbine. Thus, one could say that using such an APU is using a small gas turbine to start a larger one. High pressure air from the compressor section of the APU is bled off through a system of pipes to the engines where it is directed into the starting system. This "bleed air" is directed into a mechanism to start the engine turning and begin pulling in air. When the rotating speed of the engine is sufficient to pull in enough air to support combustion, fuel is introduced and ignited. Once the engine ignites and reaches idle speed, the bleed air is shut off.

The APUs on aircraft such as the Boeing 737 and Airbus A320 can be seen at the extreme rear of the aircraft. This is the typical location for an APU on most commercial airliners although some may be within the wing root (Boeing 727) or the aft fuselage (DC-9/ MD80) as examples and some military transports carry their APU's in one of the main landing gear pods (C-141).

The APUs also provide enough power to keep the cabin lights, pressure and other systems on while the engines are off. The valves used to control the airflow are usually electrically controlled. They automatically close at a pre-determined speed. As part of the starting sequence on some engines fuel is combined with the supplied air and burned instead of using just air. This usually produces more power per unit weight.

Usually an APU is started by its own electric starter motor which is switched off at the proper speed automatically. When the main engine starts up and reaches the right conditions, this auxiliary unit is then switched off and disengages slowly.

Hydraulic pumps can also be used to start some engines through gears. The pumps are electrically controlled on the ground.

A variation of this is the APU installed in a Boeing F/A-18 Hornet; it is started by a hydraulic motor, which itself receives energy stored in an accumulator. This accumulator is recharged after the right engine is started and develops hydraulic pressure, or by a hand pump in the right hand main landing gear well.

Ignition

Usually there are 2 igniter plugs in different positions in the combustion system. A high voltage spark is used to ignite the gases. The voltage is stored up from a low voltage supply provided by the starter system. It builds up to the right value and is then released as a high energy spark. Depending on various conditions, the igniter continues to provide sparks to prevent combustion from failing if the flame inside goes out. Of course, in the event that the flame does go out, there must be provision to relight. There is a limit of altitude and air speed at which an engine can obtain a satisfactory relight.

For example, the General Electric F404-400 uses one ignitor for the combustor and one for the afterburner; the ignition system for the A/B incorporates an ultraviolet flame sensor to activate the ignitor.

It should be noted that most modern ignition systems provide enough energy to be a lethal hazard should a person be in contact with the electrical lead when the system is activated, so team communication is vital when working on these systems.

Lubrication system

A lubrication system serves to ensure lubrication of the bearings and to maintain sufficiently cool temperatures, mostly by eliminating friction.

The lubrication system as a whole should be able to prevent foreign material from entering the plane, and reaching the bearings, gears, and other moving parts. The lubricant must be able to flow easily at relatively low temperatures and not disintegrate or break down at very high temperatures.

Usually the lubrication system has subsystems that deal individually with the pressure of an engine, scavenging, and a breather.

The pressure system components are an oil tank and de-aerator*,* main oil pump*,* main oil filter/filter bypass valve*,* pressure regulating valve (PRV)*,* oil cooler/by pass valve *and* tubing/jets.
Usually the flow is from the tank to the pump inlet and PRV, pumped to main oil filter or its bypass valve and oil cooler, then through some more filters to jets in

the bearings.

Using the PRV method of control, means that the pressure of the feed oil must be below a critical value (usually controlled by other valves which can leak out excess oil back to tank if it exceeds the critical value). The valve opens at a certain pressure and oil is kept moving at a constant rate into the bearing chamber.

If the engine speed increases, the pressure within the bearing chamber also increases, which means the pressure difference between the lubricant feed and the chamber reduces which could reduce slow rate of oil when it is needed even more. As a result, some PRVs can adjust their spring force values using this pressure change in the bearing chamber proportionally to keep the lubricant flow constant.

Advanced designs

J-58 combined ramjet/turbojet

The SR-71's Pratt & Whitney J58 engines were rather unusual. They could convert in flight from being largely a turbojet to being largely a compressor-assisted ramjet. At high speeds (above Mach 2.4), the engine used variable geometry vanes to direct excess air through 6 bypass pipes from downstream of the fourth compressor stage into the afterburner. 80% of the SR-71's thrust at high speed was generated in this way, giving much higher thrust, improving specific impulse by 10-15%, and permitting continuous operation at Mach 3.2. The name coined for this setup is *turbo-ramjet*.

Hydrogen fuelled jet engines

Jet engines can be run on almost any fuel. Hydrogen is a highly desirable fuel, as, although the energy per mole is not unusually high, the molecule is very much lighter than other molecules. It turns out that the energy per kg of hydrogen is twice that of more common fuels and this gives twice the specific impulse. In addition jet engines running on hydrogen are quite easy to build- the first ever turbojet was run on hydrogen.

However, in almost every other way, hydrogen is problematic. The downside of hydrogen is its density, in gaseous form the tanks are impractical for flight, but even in liquid form it has a density one fourteenth that of water. It is also deeply cryogenic and requires very significant insulation that precludes it being stored in wings. The overall vehicle ends up very large, and they would be difficult for most airports to accommodate. Finally, pure hydrogen is not found in nature, and must be manufactured either via steam reforming or expensive electrolysis. Both are relatively inefficient processes.

Precooled jet engines

An idea originated by Robert P. Carmichael in 1955 is that hydrogen fuelled engines could theoretically have much higher performance than hydrocarbon fuelled engines if a heat exchanger were used to cool the incoming air. The low temperature allows lighter materials to be used, a higher mass-flow through the engines, and permits combustors to inject more fuel without overheating the engine.

This idea leads to plausible designs like SABRE, that might permit single-stage-to-orbit, and ATREX, that might permit jet engines to be used up to hypersonic speeds and high altitudes for boosters for launch vehicles. The idea is also being researched by the EU for a concept to achieve non-stop antipodal supersonic passenger travel at Mach 5 (Reaction Engines A2).

Nuclear-powered ramjet

Project Pluto was a nuclear-powered ramjet, intended for use in a cruise missile. Rather than combusting fuel as in regular jet engines, air was heated using a high-temperature, unshielded nuclear reactor. This dramatically increased the engine burn time, and the ramjet was predicted to be able to cover any required distance at supersonic speeds (Mach 3 at tree-top height).

However, there was no obvious way to stop it once it had taken off, which would be a great disadvantage in any non-disposable application. Also, because the reactor was unshielded, it was dangerous to be in or around the flight path of the vehicle (although the exhaust itself wasn't radioactive). These disadvantages limit the application to warhead delivery system for all-out nuclear war, which it was being designed for.

Scramjets

Scramjets are an evolution of ramjets that are able to operate at much higher speeds than any other kind of airbreathing engine. They share a similar structure with ramjets, being a specially-shaped tube that compresses air with no moving parts through ram-air compression. Scramjets, however, operate with supersonic airflow through the entire engine. Thus, scramjets do not have the diffuser required by ramjets to slow the incoming airflow to subsonic speeds.

Scramjets start working at speeds of at least Mach 4, and have a maximum useful speed of approximately Mach 17. Due to aerodynamic heating at these high speeds, cooling poses a challenge to engineers.

Environmental considerations

Jet engines are usually run on fossil fuel propellant, and in that case, are a net source of carbon to the atmosphere.

Some scientists believe that jet engines are also a source of global dimming due to the water vapour in the exhaust causing cloud formations.

Nitrogen compounds are also formed from the combustion process from atmospheric nitrogen. At low altitudes this is not thought to be especially harmful, but for supersonic aircraft that fly in the stratosphere some destruction of ozone may occur.

Sulphates are also emitted if the fuel contains sulphur.

Safety and reliability

Jet engines are usually very reliable and have a very good safety record. However failures do sometimes occur.

One class of failures that has caused accidents in particular is uncontained failures, where rotary parts of the engine break off and exit through the case. These can cut fuel or control lines, and can penetrate the cabin. Although fuel and control lines are usually duplicated for reliability the United Airlines Flight 232 was caused when all control lines were simultaneously severed.

The most likely failure is compressor blade failure, and modern jet engines are designed with structures that can catch these blades and keep them contained them within the engine casing. Verification of a jet engine design involves testing that this system works correctly.

Bird strike

Bird strike is an aviation term for a collision between a bird and an aircraft. It is a common threat to aircraft safety and has caused a number of fatal accidents. In 1988 an Ethiopian Airlines Boeing 737 sucked pigeons into both engines during take-off and then crashed in an attempt to return to the Bahir Dar airport; of the 104 people aboard, 35 died and 21 were injured. In another incident in 1995, a Dassault Falcon 20 crashed at a Paris airport during an emergency landing attempt after sucking lapwings into an engine, which caused an engine failure and a fire in the airplane fuselage; all 10 people on board were killed.

Modern jet engines have the capability of surviving an ingestion of a bird. Small fast planes, such as military jet fighters, are at higher risk than big heavy multi-engine ones. This is due to the fact that the fan of a high-bypass turbofan engine, typical on transport aircraft, acts as a centrifugal separator to force ingested materials (birds, ice, etc.) to the outside of the fan's disc. As a result, such materials go through the relatively unobstructed bypass duct, rather than through the core of the engine, which contains the smaller and more delicate compressor blades. Military aircraft designed for high-speed flight typically have pure turbojet, or low-bypass turbofan engines, increasing the risk that ingested materials will get into the core of the engine to cause damage.

The highest risk of the bird strike is during the takeoff and landing, in low altitudes, which is in the vicinity of the airports.

Longship

2008/9 Schools Wikipedia Selection. Related subjects: Air & Sea transport; British History 1500 and before (including Roman Britain)

The Oseberg longship (Viking Ship Museum, Norway)

Oseberg longship from the front, one of the most stunning expressions of Norse art and craftsmanship

Longships were ships primarily used by the Scandinavian Vikings and the Saxons to raid coastal and inland settlements during the European Middle Ages. They are often called " longboats", but "longship" is more accurate. The vessels were also used for long distance trade and commerce, and for exploratory voyages to Iceland, Greenland, and beyond. Longship design evolved over several centuries and was fully developed by about the 9th century. In Norway traditional longships were used until the 13th century, and the character and appearance of these ships were reflected in western Norwegian boat-building traditions until the early 20th century.

The longship was characterized as a graceful, long, narrow, light wooden boat with a shallow draft designed for speed. The ship's shallow draft allowed navigation in waters only one metre deep and permitted rapid beach landings, while its light weight enabled it to be carried over portages. Longships were also symmetrical, allowing the ship to reverse direction quickly. Longships were fitted with oars along almost the entire length of the boat itself. Later versions sported a rectangular sail on a single mast which was used to augment the effort of the rowers, particularly during long journeys.

Longships were the epitome of Scandinavian naval power at the time, and were highly valued possessions. They were often owned by coastal farmers and commissioned by the king in times of conflict, in order to build a powerful naval force. While longships were used by the Vikings in warfare, they were troop transports, not warships. In the tenth century, these boats would sometimes be tied together in battle to form a steady platform for infantry warfare.They were called dragonships by enemies such as the English.

Development history

The famous Viking longships did not suddenly spring into being, but developed over hundreds of years. Archaeologists have uncovered a number of ships and boats showing this development, and rock carvings and runestones which predate the longships also indicate a long shipbuilding tradition in Scandinavia.

Early ships

The Hjortspring ship

One of the early precursors of the longship was the Hjortspring boat. This 13 m (40 feet) boat was found on the Hjortspring farm on the Danish island of Als. It was probably built between 200 B.C and 350 B.C., of five limewood planks. The boat has been interpreted as an early war canoe that was lowered into a pool as a sacrifice. Its design already shows some of the features of later longships, such as clinker construction. The boards for the hull were cut into wedge-shaped pieces or "cleats", and hazelwood ribs were fastened inside. The method of attaching the boards and gunwales was later adapted to longships to make them flexible for ocean voyages. The boat was propelled by paddles.

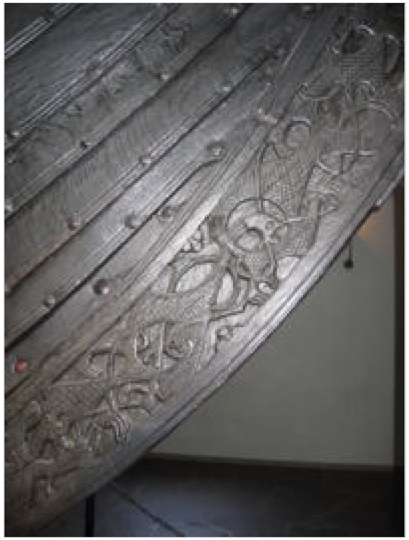
The Nydam ship

The Nydam ship (or *Nydam Oak Boat*) had a much improved design compared to the Hjortspring boat. It was one of three ships found in a series of excavations in the middle of the 19th century 8km from Sønderborg near Schleswig on the German-Danish border. The ship was dated to about 315 A.D using dendochronology. The Nydam ship was both larger and much more technologically advanced than the Hjortspring boat. The ship measured 23 m (75 feet) in length and was built from oak. It was originally believed that its planks (technically " strakes") were of a single piece running the full length of the hull. However, while sampling the wood for dating, it was discovered that they were composed of a few long pieces carefully connected by invisible joints. The planks were held together by iron rivets and formed a curved prow and stern. The Nydam ship is the first known ship in Northern Europe to use oars rather than paddles for propulsion. The oars were held in place by bent branches secured to the rail. This allowed greater speed and easier rowing for the crew. The ship had a narrow, V-shaped hull giving it superior speed and agility. However, this also made it to rather unstable, and unable to support mast and sail.

The Kvalsund ship

Two ships dated to the 7th century were found in Kvalsund, Norway. Both were of similar design. Despite the shorter length of about 18m (61 feet), the larger Kvalsund ship was far wider than the older vessels mentioned above, with a width of about 3.5 m (10 feet). It had a clearly defined, strong keel. These key improvements allowed it to maintain a course even under adverse weather conditions. The Kvalsund ship had oars that were fastened to the rails with wooden pegs or trenails. No sailing rig has been found, although the ship certainly could have carried mast and sail. Apart from the lack of rigging, the Kvalsund ship already had most of the characteristics of a true longship.

The Oseberg ship


Detail from the Oseberg ship

The continuing evolution of the Scandinavian sailing ships is evident in the Oseberg ship, which has been dated to c. 815–820 A.D. and was found in a burial mound in Vestfold south of Oslo, Norway. The Oseberg can be considered one of the first true longships. It features a built-in mast and mast partner. Compared to later ships, the Oseberg is a rather frail vessel, and it is thought that it was only used in coastal waters, or built especially for the funeral.

By this period the distance between the ship's ribs was a standard length and the ribs were stronger. The hulls had more of a V shape and the length expanded from gunwale to gunwale. These new hulls had poor lateral stability but made up for it in speed. It had wood fastened together instead of single pieces which allowed greater stability and agility.

The Oseberg has a length of 21.5 meters, a width of 5 meters and a total weight of 11 tons.

The Vikings were excellent sailors their boats were called longships. Longships were light, sleek, stable, strong and easy to manoeuvre. Being long and thin they made great warships. The hull cut through the water fast. The boat was also flexible, so it moved with the action of the waves. The longship has a flat bottom with a shallow draught which allowed the Viking to sail into shallow waters bays and even the shore line. Longships are around 28 – 30 meters long in size and built to hold more than 100 men. The boats speed can get up to 30 – 35 kilometres per hour because the Vikings had both oars and sails so they could keep going in any weather condition.

They were constructed out of raw timber. The kneel (the bottom of the boat) was made out of a single trunk, planks were made from split timber, sternposts are cut from large curved logs, angled sections were cut from strong branches and curved sections were cut from curved branches.

In building a longship you would use up to 10 or more tools.

Types of longship

Longships can be classified into a number of different types, depending on size, construction details, and prestige.

Snekke (snekkja)

The snekke was the smallest vessel that would still be considered a longship. A typical snekke might have a length of 17 m, a width of 2.5 m, and a draught of only 0.5 m. It would carry a crew of about 25 men.

Snekkes were one of the most common types of ship. According to historical lore, Canute the Great used 1400 in Norway in 1028, and William the Conqueror used about 600 for the invasion of Britain in 1066.

The Norwegian snekkes, designed for deep fjords and Atlantic weather, typically had more draft than the Danish model designed for low coasts and beaches. Snekkes were so light that they had no need of ports – they could simply be beached, and potentially even carried across a portage.

The snekke continued to evolve after the end of the Viking age, with later Norwegian examples becoming larger and heavier than Viking age ships.

Dragon ships

Dragon ships are known from historical sources, such as the 13th century *Göngu-Hrólfs Saga* (the Saga of Rollo). Here, the ships are described as elegant and ornately decorated, and used by those who went *í Viking* (raiding and plundering). According to the historical sources the ships' prows carried carvings of menacing beasts, such as dragons and snakes, allegedly to protect the ship and crew, and to ward off the terrible sea monsters of Norse mythology. It is however likely that the carvings, like those on the Oseberg ship, might have had a ritual purpose, or that the purported effect was to frighten enemies and townspeople. No true dragon ship, as defined by the sagas, has been found by archaeological excavation. Therefore, their existence is only supported by the historical sources.

Roskilde ships

The **largest** longships so far found, were discovered by Danish archaeologists in Roskilde during development in the harbour-area in 1962 and 1996/7. The ship discovered in 1962, *Skuldelev 2* is an oak-built vessel possibly of the *skeid* type. It was built in the Dublin area around 1042. *Skuldelev 2* could carry a crew of some 70-80 and measures just under 100 feet (30 m) in length. In 1996/7 archaeologists discovered the remains of another ship in the harbour. This ship, called the *Roskilde 6*, has not yet been fully investigated and full details are not available. It is however thought to be around 36m long, and has been dated to the mid-11th century.

The discovery of these ships overturned the skepticism of some historians that longships of this size had ever been constructed. The Roskilde longships may have been a specialized type of cargo ship that the Vikings used for trade.

Construction

After several centuries of evolution, the fully developed longship emerged some time in the middle of the ninth century. Its long, graceful, menacing head figure carved in the stern echoed the designs of its predecessors. The mast was now squared and located toward the middle of the ship, and could be lowered and raised. The hull's sides were fastened together to allow it to flex with the waves, ensuring stability and integrity. The ships were large enough to carry cargo and passengers on long ocean voyages but still maintained speed and agility, making the longship a versatile warship and cargo carrier.

The Gokstad ship, on display at the Viking ship museum in Oslo, Norway.

Selection of wood

Wood was the fundamental material of the longship: it was used in every part of the ship, from the planks for the hull to the mast and oars. The Vikings had developed the selection and cutting of wood to a fine science. They made planks by splitting huge oak trees. The trunks were cut radially from tall trees, which contained few knots. The planks had exceptional strength, due to the fact that they were cut following the grain of the wood. The planks also were cut in such a way that they did not shrink or warp as they dried. Shipbuilders used fresh-cut trees rather than seasoned timber because it was easier to work. Curved pieces were made from trees that had grown naturally in that shape. This allowed the part to be made from a single piece of wood, cutting down the weight of the ship. About 100 oak trees were used to build a longship.

Keel, stems and hull

The Viking shipbuilders had no written diagrams or standard written design plan. The shipbuilder pictured the longship before its construction, and the ship was then built from the ground up. The keel and stems were made first. The shape of the stem was based on segments of circles of varying sizes. The next step was building the strakes – the lines of planks joined endwise from stern to stern. Nearly all longships were clinker built, meaning that each hull plank overlapped the next.

As the strakes reached the desired height, the interior frame and cross beams were added. The parts were held together with iron rivets, as well as spruce strips that were fastened to the ribs inside of the keel. Longships had about five rivets for each yard of plank.

The longships' wider hulls provided strength beneath the waterline which gave more stability, making the longship less likely to tip or bring in water. The hull was waterproofed with moss drenched in tar. In the autumn the ships would be tarred and then left in a boathouse over the winter to allow time for the tar to dry. To keep the sea out, wooden disks were put into the oar holes. These could be shut from the inside when the oars were not in use.

Sail and mast

Even though no longship sail has been found, accounts verify that longships had square sails. Sails measured perhaps 35 to 40 feet across, and were made of wadmill (rough wool) which was woven by looms. Unlike the knarrs, the longship sail was not stitched.

The sail was held in place by the mast. The mast was supported by a large block of wood called "kerling" ("Old Woman" in Old Norse). (Trent) The kerling was made of oak, and was as tall as a Viking man. The kerling lay across the two ribs and ran width-wise along the keel. The kerling also had a companion: the "mast fish", a wooden piece above the kerling that provided extra help in keeping the mast erect. (information need for how long boat construction and sail creation needed.)

Navigation and propulsion

Navigation

A replica of the Gokstad ship, named *Viking* was sailed across the Atlantic to the World's Columbian Exposition in 1893

The Vikings were experts in judging speed and wind direction, and in knowing the current and when to expect high and low tides. Viking navigational techniques are not well understood, but historians postulate that the Vikings probably had some sort of primitive astrolabe and used the stars to plot their course.

A Viking named Stjerner Oddi compiled a chart showing the direction of dawn and twilight, which enabled navigators to sail longships from place to place with ease. Almgren, an earlier Viking, told of another method: "All the measurements of angles were made with what was called a 'half wheel' (a kind of half sun-diameter which corresponds to about sixteen seconds of arc). This was something that was known to every skipper at that time, or to the long-voyage pilot or 'kendtmand' ('man who knows the way') who sometimes went along on voyages... When the sun was in the sky, it was not, therefore, difficult to find the four points of the compass, and determining latitude did not cause any problems either." (Algrem)

Birds provided a helpful guide to finding land. A Viking legend states that Vikings used to take caged crows aboard ships and let them loose if they got lost. The crows would instinctively find land, giving the Viking navigators their direction. Little is known of Viking compasses, though Viking legends do tell of small magnetic stones floating on a piece of wood in water to provide a point of navigational reference.

Propulsion

The longship had two methods of propulsion: oars and sail. At sea, the sail enabled longships to travel faster than by oar and to cover long distances. Sails could be raised or lowered quickly. Oars were used when land was spotted, to gain speed quickly (when there was no wind), and to get the boat started. In combat, the variability of wind power made rowing the chief means of propulsion.

Longships were not fitted with benches. When rowing, the crew sat on sea chests (chests containing their personal possessions) that would otherwise take up space. The chests were made the same size and were the perfect height for a Viking to sit on and row. Longships had hooks for oars to fit into, but smaller oars were also used, with crooks or bends to be used as oarlocks. If there were no holes then a loop of rope kept the oars in place.

Legacy

The Vikings were major contributors to the shipbuilding technology of their day. Their shipbuilding methods spread through extensive contact with other cultures, and ships from the 11th and 12th centuries are known to borrow many of the longships' design features, despite the passing of many centuries. The 'Lancha Poveira', a boat from Póvoa de Varzim, Portugal, originates from the longship, but without a long stern and bow, and with a Mediterranean sail. It was used until the 1950s. Today there is just one boat: *Fé em Deus*.


The 'Lancha Poveira', a boat from Póvoa de Varzim, Portugal.

Many historians, archaeologists and adventurers have reconstructed longships in an attempt to understand how they worked. These re-creators have been able to identify many of the advances that the Vikings implemented in order to make the longship a superior vessel. One replica longship covered 223 nautical miles in a single day, and another re-creator was able to go faster than 8 knots in his longship.

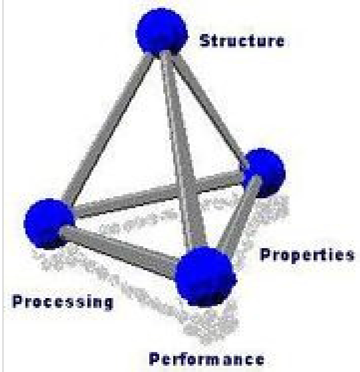
The longship was a master of all trades: it was wide and stable, yet light, fast and nimble. With all these qualities combined in one ship, the longship was unrivaled for centuries, until the arrival of the great gunboats and galleons.

Famous longships

- The Oseberg ship and the Gokstad ship of Oslo.
- The Ormen Lange ("*The Long Serpent*") was the most famous longship of Norwegian king Olaf Tryggvason.
- The Mora was the ship given to William the Conqueror by his wife, Matilda, and used as the flagship in the conquest of England.
- The Sea Stallion, the largest Viking ship replica ever made, is a new 30 meter replica of the skuldelev 2, and is going to be sailed from Roskilde, Denmark to Dublin in 2007 to commemorate the voyage of the original.

Materials science

2008/9 Schools Wikipedia Selection. Related subjects: Engineering



The Materials Science Tetrahedron, which often also includes Characterization at the centre

Materials science or **materials engineering** is an interdisciplinary field involving the properties of matter and its applications to various areas of science and engineering. This science investigates the relationship between the structure of materials and their properties. It includes elements of applied physics and chemistry, as well as chemical, mechanical, civil and electrical engineering. With significant media attention to nanoscience and nanotechnology in recent years, materials science has been propelled to the forefront at many universities. It is also an important part of forensic engineering and forensic materials engineering, the study of failed products and components.

History

The material of choice of a given era is often its defining point; the Stone Age, Bronze Age, and Steel Age are examples of this. Materials science is one of the oldest forms of engineering and applied science, deriving from the manufacture of ceramics. Modern materials science evolved directly from metallurgy, which itself evolved from mining. A major breakthrough in the understanding of materials occurred in the late 19th century, when Willard Gibbs demonstrated that thermodynamic properties relating to atomic structure in various phases are related to the physical properties of a material. Important elements of modern materials science are a product of the space race: the understanding and engineering of the metallic alloys, and silica and carbon materials, used in the construction of space vehicles enabling the exploration of space. Materials science has driven, and been driven by, the development of revolutionary technologies such as plastics, semiconductors, and biomaterials.

Before the 1960s (and in some cases decades after), many *materials science* departments were named *metallurgy* departments, from a 19th and early 20th century emphasis on metals. The field has since broadened to include every class of materials, including: ceramics, polymers, semiconductors, magnetic materials, medical implant materials and biological materials.

Fundamentals of materials science

In materials science, rather than haphazardly looking for and discovering materials and exploiting their properties, one instead aims to understand materials fundamentally so that new materials with the desired properties can be created.

The basis of all materials science involves relating the desired properties and relative performance of a material in a certain application to the structure of the atoms and phases in that material through characterization. The major determinants of the structure of a material and thus of its properties are its constituent chemical elements and the way in which it has been processed into its final form. These, taken together and related through the laws of thermodynamics, govern a material's microstructure, and thus its properties.

An old adage in materials science says: "materials are like people; it is the defects that make them interesting". The manufacture of a perfect crystal of a material is currently physically impossible. Instead materials scientists manipulate the defects in crystalline materials such as precipitates, grain boundaries (Hall-Petch relationship), interstitial atoms, vacancies or substitutional atoms, to create materials with the desired properties.

Not all materials have a regular crystal structure. Polymers display varying degrees of crystallinity, and many are completely non-crystalline. Glasses, some ceramics, and many natural materials are amorphous, not possessing any long-range order in their atomic arrangements. The study of polymers combines elements of chemical and statistical thermodynamics to give thermodynamic, as well as mechanical, descriptions of physical properties.

In addition to industrial interest, materials science has gradually developed into a field which provides tests for condensed matter or solid state theories. New physics emerge because of the diverse new material properties which need to be explained.

Materials in industry

Radical materials advances can drive the creation of new products or even new industries, but stable industries also employ materials scientists to make incremental improvements and troubleshoot issues with currently used materials. Industrial applications of materials science include materials design, cost-benefit tradeoffs in industrial production of materials, processing techniques (casting, rolling, welding, ion implantation, crystal growth, thin-film deposition, sintering, glassblowing, etc.), and analytical techniques (characterization techniques such as electron microscopy, x-ray diffraction, calorimetry, nuclear microscopy (HEFIB), Rutherford backscattering, neutron diffraction, etc.).

Besides material characterisation, the material scientist/engineer also deals with the extraction of materials and their conversion into useful forms. Thus ingot casting, foundry techniques, blast furnace extraction, and electrolytic extraction are all part of the required knowledge of a metallurgist/engineer. Often the presence, absence or variation of minute quantities of secondary elements and compounds in a bulk material will have a great impact on the final properties of the materials produced, for instance, steels are classified based on 1/10th and 1/100 weight percentages of the carbon and other alloying elements they contain. Thus, the extraction and purification techniques employed in the extraction of iron in the blast furnace will have an impact of the quality of steel that may be produced.

The overlap between physics and materials science has led to the offshoot field of *materials physics*, which is concerned with the physical properties of materials. The approach is generally more macroscopic and applied than in condensed matter physics. See important publications in materials physics for more details on this field of study.

The study of metal alloys is a significant part of materials science. Of all the metallic alloys in use today, the alloys of iron (steel, stainless steel, cast iron, tool steel, alloy steels) make up the largest proportion both by quantity and commercial value. Iron alloyed with various proportions of carbon gives low, mid and high carbon steels. For the steels, the hardness and tensile strength of the steel is directly related to the amount of carbon present, with increasing carbon levels also leading to lower ductility and toughness. The addition of silicon and graphitization will produce cast irons (although some cast irons are made precisely with no graphitization). The addition of chromium, nickel and molybdenum to carbon steels (more than 10%) gives us stainless steels.

Other significant metallic alloys are those of aluminium, titanium, copper and magnesium. Copper alloys have been known for a long time (since the Bronze Age), while the alloys of the other three metals have been relatively recently developed. Due to the chemical reactivity of these metals, the electrolytic extraction processes required were only developed relatively recently. The alloys of aluminium, titanium and magnesium are also known and valued for their high strength-to-weight ratios and, in the case of magnesium, their ability to provide electromagnetic shielding. These materials are ideal for situations where high strength-to-weight ratios are more important than bulk cost, such as in the aerospace industry and certain automotive engineering applications.

Other than metals, polymers and ceramics are also an important part of materials science. Polymers are the raw materials (the resins) used to make what we commonly call plastics. Plastics are really the final product, created after one or more polymers or additives have been added to a resin during processing, which is then shaped into a final form. Polymers which have been around, and which are in current widespread use, include polyethylene, polypropylene, PVC, polystyrene, nylons, polyesters, acrylics, polyurethanes, and polycarbonates. Plastics are generally classified as "commodity", "specialty" and "engineering" plastics.

PVC (polyvinyl-chloride) is widely used, inexpensive, and annual production quantities are large. It lends itself to an incredible array of applications, from artificial leather to electrical insulation and cabling, packaging and containers. Its fabrication and processing are simple and well-established. The versatility of PVC is due to the wide range of plasticisers and other additives that it accepts. The term "additives" in polymer science refers to the chemicals and compounds added to the polymer base to modify its material properties.

Polycarbonate would be normally considered an engineering plastic (other examples include PEEK, ABS). Engineering plastics are valued for their superior strengths and other special material properties. They are usually not used for disposable applications, unlike commodity plastics.

Specialty plastics are materials with unique characteristics, such as ultra-high strength, electrical conductivity, electro-fluorescence, high thermal stability, etc.

It should be noted here that the dividing line between the various types of plastics is not based on material but rather on their properties and applications. For instance, polyethylene (PE) is a cheap, low friction polymer commonly used to make disposable shopping bags and trash bags, and is considered a commodity plastic, whereas Medium-Density Polyethylene MDPE is used for underground gas and water pipes, and another variety called Ultra-high Molecular Weight Polyethylene UHMWPE is an engineering plastic which is used extensively as the glide rails for industrial equipment and the low-friction socket in implanted hip joints.

Another application of material science in industry is the making of composite materials. Composite materials are structured materials composed of two or more macroscopic phases. An example would be steel-reinforced concrete; another can be seen in the "plastic" casings of television sets, cell-phones and so on. These plastic casings are usually a composite material made up of a thermoplastic matrix such as acrylonitrile-butadiene-styrene (ABS) in which calcium carbonate chalk, talc, glass fibres or carbon fibres have been added for added strength, bulk, or electro-static dispersion. These additions may be referred to as reinforcing fibres, or dispersants, depending on their purpose.

Classes of materials (by bond types)

Materials science encompasses various classes of materials, each of which may constitute a separate field. Materials are sometimes classified by the type of bonding present between the atoms:

1. Ionic crystals
2. Covalent crystals
3. Metals
4. Intermetallics
5. Semiconductors
6. Polymers
7. Composite materials
8. Vitreous materials

Sub-fields of materials science

- Nanotechnology – rigorously, the study of materials where the effects of quantum confinement, the Gibbs-Thomson effect, or any other effect only present at the nanoscale is the defining property of the material; but more commonly, it is the creation and study of materials whose defining structural properties are anywhere from less than a nanometer to one hundred nanometers in scale, such as molecularly engineered materials.
- Microtechnology - study of materials and processes and their interaction, allowing microfabrication of structures of micrometric dimensions, such as MicroElectroMechanical Systems (MEMS).
- Crystallography – the study of how atoms in a solid fill space, the defects associated with crystal structures such as grain boundaries and dislocations, and the characterization of these structures and their relation to physical properties.
- Materials Characterization – such as diffraction with x-rays, electrons, or neutrons, and various forms of spectroscopy and chemical analysis such as Raman spectroscopy, energy-dispersive spectroscopy (EDS), chromatography, thermal analysis, electron microscope analysis, etc., in order to understand and define the properties of materials. See also List of surface analysis methods
- Metallurgy – the study of metals and their alloys, including their extraction, microstructure and processing.
- Biomaterials – materials that are derived from and/or used with biological systems.
- Electronic and magnetic materials – materials such as semiconductors used to create integrated circuits, storage media, sensors, and other devices.
- Tribology – the study of the wear of materials due to friction and other factors.
- Surface science/Catalysis – interactions and structures between solid-gas solid-liquid or solid-solid interfaces.
- Ceramography – the study of the microstructures of high-temperature materials and refractories, including structural ceramics such as RCC, polycrystalline silicon carbide and transformation toughened ceramics

Some practitioners often consider rheology a sub-field of materials science, because it can cover any material that flows. However, modern rheology typically deals with non-Newtonian fluid dynamics, so it is often considered a sub-field of continuum mechanics. See also granular material.

- Glass Science – any non-crystalline material including inorganic glasses, vitreous metals and non-oxide glasses.
- Forensic engineering – the study of how products fail, and the vital role of the materials of construction
- Forensic materials engineering – the study of material failure, and the light it sheds on how engineers specify materials in their product

Topics that form the basis of materials science

- Thermodynamics, statistical mechanics, kinetics and physical chemistry, for phase stability, transformations (physical and chemical) and diagrams.
- Crystallography and chemical bonding, for understanding how atoms in a material are arranged.
- Mechanics, to understand the mechanical properties of materials and their structural applications.
- Solid-state physics and quantum mechanics, for the understanding of the electronic, thermal, magnetic, chemical, structural and optical properties of materials.
- Diffraction and wave mechanics, for the characterization of materials.
- Chemistry and polymer science, for the understanding of plastics, colloids, ceramics, liquid crystals, solid state chemistry, and polymers.
- Biology, for the integration of materials into biological systems.
- Continuum mechanics and statistics, for the study of fluid flows and ensemble systems.
- Mechanics of materials, for the study of the relation between the mechanical behaviour of materials and their microstructures.

Important Journals

- Chemistry of Materials
- Nature Materials
- Acta Materialia
- JOM
- Advanced Materials
- Computational materials science
- Advanced Functional Materials
- Journal of Materials Chemistry
- Journal of Materials Online - Open Access
- Metallurgical and Materials Transactions
- Journal of Materials Research
- Journal of Materials Science
- Federation of European Materials Science Societies Newsletter
- AMMTIAC eNews/Quarterly Advanced materials, manufacturing, and testing. (Free subscription)

Mechanical engineering

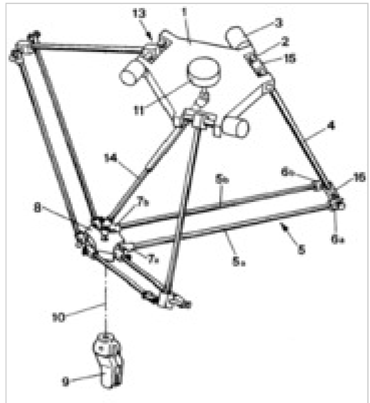
2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Mechanical Engineering is an engineering discipline that involves the application of principles of physics for analysis, design, manufacturing, and maintenance of mechanical systems. It requires a solid understanding of key concepts including mechanics, kinematics, thermodynamics and energy. Mechanical engineers use these principles and others in the design and analysis of motor vehicles, aircraft, heating & cooling systems, watercraft, manufacturing plants, industrial equipment and machinery, medical devices and more.

Development

Mechanical engineers design and build engines and power plants...

...structures and vehicles of all sizes...



...and moving mechanisms, machines, and robots.

Mechanical engineering could be found in many ancient and medieval societies throughout the globe. In ancient Greece, the works of Archimedes (287 BC–212 BC), and Heron of Alexandria (10–70 AD) deeply influenced mechanics in the Western tradition. In ancient China, there were also many notable figures, such as Zhang Heng (78–139 AD) and Ma Jun (200–265 AD). The medieval Chinese horologist and engineer Su Song (1020–1101 AD) incorporated an escapement mechanism into his astronomical clock tower two centuries before any escapement could be found in clocks of medieval Europe, as well as the world's first known endless power-transmitting chain drive.

During the early 19th century in Britain, the development of machine-tools led mechanical engineering to develop as a separate field within engineering, providing manufacturing machines and the engines to power them. The first British professional society of mechanical engineers was formed in 1847, making mechanical engineering the second-oldest branch of engineering behind civil, formed 30 years earlier. In the United States, the first mechanical engineering professional society was formed in 1880, making it the third oldest type of engineering behind civil (1852) and mining & metallurgical (1871). The first schools in the United States to offer an engineering education were the United States Military Academy in 1817, an institution now known as Norwich University in 1819, and Rensselaer Polytechnic Institute in 1825. Education in mechanical engineering has historically been based on a strong foundation in mathematics and science; this is followed by courses emphasizing the application of this knowledge and studies in the social sciences and humanities to give the engineer a broader education.

Education

Degrees in mechanical engineering are offered at universities in most industrialized nations. In China, India, North America and elsewhere, mechanical engineering programs typically take 4 to 5 years and result in a Bachelor of Science (BSc), Bachelor of Technology (BTech), Bachelor of Engineering (B.Eng), or a Bachelor of Applied Science (B.A.Sc.) with emphasis in mechanical engineering. In Spain, Portugal and most of South America where the (BSc) or (BTech) programs have not been adopted, the formal name for the degree is "Mechanical Engineer" and the course work is based on 5–6 years of training.

In the U.S., most undergraduate Mechanical Engineering programs are accredited by the Accreditation Board for Engineering and Technology (ABET) to ensure similar course requirements and standards between universities. The ABET web site lists 276 accredited Mechanical Engineering programs as of June 19, 2006. Mechanical Engineering programs in Canada are accredited by the Canadian Engineering Accreditation Board (CEAB), and most other countries offering engineering degrees have similar accreditation societies.

Some Mechanical Engineers go on to pursue a postgraduate degree such as a Master of Engineering, Master of Science, Master of Engineering Management (MEng.Mgt, MEM), a Doctor of Philosophy in Engineering (EngD, PhD) or an Engineer's degree. The Master's and Engineer's degrees may consist of either research, coursework or a mixture of the two. The Doctor of Philosophy consists of a significant research component and is often viewed as the entry point to academia.

Coursework

Mechanical engineering programs generally cover the same fundamental subjects, to meet standards set by each country's accreditation society. Engineering programs in the U.S., for instance, are required by ABET to show that their students can "work professionally in both thermal and mechanical systems areas." This is to ensure a minimum level of competence among graduating engineers and to maintain confidence in the engineering profession as a whole. The specific courses required to graduate, however, may differ from program to program. Universities will often combine multiple subjects into a single class or split a subject into multiple classes, depending on the faculty available and the university's major area(s) of research. Fundamental subjects of mechanical engineering usually include:

- statics & dynamics
- strength of materials & solid mechanics
- instrumentation and measurement
- thermodynamics, heat transfer, energy conversion, and HVAC
- fluid mechanics & fluid dynamics
- mechanism design (including kinematics and dynamics)
- manufacturing technology or processes
- hydraulics & pneumatics
- engineering design
- mechatronics and control theory
- drafting, CAD (usually including Solid modeling), and CAM

Mechanical engineers are also expected to understand and be able to apply basic concepts from chemistry, chemical engineering, electrical engineering, civil engineering, and physics. Most mechanical engineering programs include several semesters of calculus, as well as advanced mathematical concepts which may include differential equations and partial differential equations, linear and modern algebra, and differential geometry, among others.

In addition to the core mechanical engineering curriculum, many mechanical engineering programs offer more specialized programs and classes, such as mechatronics / robotics, transport and logistics, cryogenics, fuel technology, automotive engineering, biomechanics, vibration, optics and others, if a separate department does not exist for these subjects.

Most mechanical engineering programs also require varying amounts of research or community projects to gain practical problem-solving experience. Mechanical engineering students usually hold one or more internships while studying, though this is not typically mandated by the university.

License

Engineers may seek license by a state, provincial, or national government. The purpose of this process is to ensure that engineers possess the necessary technical knowledge, real-world experience, and knowledge of the local legal system to practice engineering at a professional level. Once certified, the engineer is given the title of *Professional Engineer* (in the United States, Canada, Japan, South Korea and South Africa), *Chartered Engineer* (in the UK, Ireland, India and Zimbabwe), *Chartered Professional Engineer* (in Australia and New Zealand) or *European Engineer* (much of the European Union). Not all mechanical engineers choose to become licensed; those that do can be distinguished as Chartered or Professional Engineers by the post-nominal title P.E., P. Eng., or C.Eng., as in: Ryan Jones, P.Eng.

In the U.S., to become a licensed Professional Engineer, an Engineer must pass the comprehensive FE (Fundamentals of Engineering) exam, work a given number of years as an *Engineering Intern (EI)* or *Engineer-in-Training (EIT)*, and finally pass the "Principles and Practice" or PE (Practicing Engineer or Professional Engineer) exam.

In the United States, the requirements and steps of this process are set forth by the National Council of Examiners for Engineering and Surveying (NCEES), a national non-profit representing all states. In the UK, current graduates require a MSc, MEng or BEng (Hons) in order to become chartered through the Institution of Mechanical Engineers.

In most modern countries, certain engineering tasks, such as the design of bridges, electric power plants, and chemical plants, must be approved by a Professional Engineer or a Chartered Engineer. "Only a licensed engineer, for instance, may prepare, sign, seal and submit engineering plans and drawings to an public authority for approval, or to seal engineering work for public and private clients." This requirement can be written into state and provincial legislation, such as Quebec's Engineer Act. In other countries, such as Australia, no such legislation exists; however, practically all certifying bodies maintain a code of ethics independent of legislation that they expect all members to abide by or risk expulsion.

Salaries and workforce statistics

The total number of engineers employed in the U.S. in 2004 was roughly 1.4 million. Of these, 226,000 were mechanical engineers (15.6%), second only to civil engineers in size at 237,000 (16.4%). The total number of mechanical engineering jobs in 2004 was projected to grow 9 to 17%, with average starting salaries being $50,236 with a bachelor's degree, $59,880 with a master's degree, and $68,299 with a doctorate degree. This places mechanical engineering at 8th of 14 among engineering bachelors degrees, 4th of 11 among masters degrees, and 6th of 7 among doctorate degrees in average annual salary. The median annual earning of mechanical engineers in the U.S. workforce is roughly $63,000. This number is highest when working for the government ($72,500), and lowest when doing general purpose machinery manufacturing in the private sector ($55,850).

Canadian engineers make an average of $29.83 per hour with 4% unemployed. The average for all occupations is $18.07 per hour with 7% unemployed. Twelve percent of these engineers are self-employed, and since 1997 the proportion of female engineers has risen to 6%.

Modern Tools

Many mechanical engineering companies, especially those in industrialized nations, have begun to incorporate Computer-aided engineering (CAE) programs into their existing design and analysis processes. For instance, companies may use Computer-aided design (CAD) for design of products in 2D and 3D. This method has many benefits, including easier and more exhaustive visualization of products, the ability to create virtual assemblies of parts, and the ease of use in designing mating interfaces and tolerances.

Other CAE programs commonly used by mechanical engineers include product lifecycle management(PLM) tools and analysis tools used to perform complex simulations. Analysis tools may be used to predict product response to expected loads, including fatigue life and manufacturability. These tools include Finite element analysis (FEA), Computational fluid dynamics (CFD), and Computer-aided manufacturing (CAM).

Using CAE programs, a mechanical design team can quickly and cheaply iterate the design process to develop a product that better meets cost, performance, and other constraints. No physical prototype need be created until the design nears completion, allowing hundreds or thousands of designs to be evaluated, instead of a relative few. In addition, CAE analysis programs can model complicated physical phenomena which cannot be solved by hand, such as viscoelasticity, complex contact between mating parts, or non-Newtonian flows

As mechanical engineering begins to merge with other disciplines, as seen in mechatronics, Multidisciplinary design optimization (MDO) is being used with other CAE programs to automate and improve the iterative design process. MDO tools wrap around existing CAE processes, allowing product evaluation to continue even after the analyst goes home for the day. They also utilize sophisticated optimization algorithms to more intelligently explore possible designs, often finding better, innovative solutions to difficult multidisciplinary design problems.

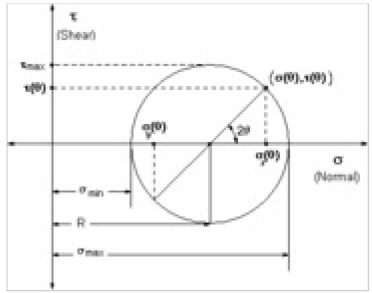
Subdisciplines

The field of mechanical engineering can be thought of as a collection of many mechanical disciplines. Several of these subdisciplines which are typically taught at the undergraduate level are listed below, with a brief explanation and the most common application of each. Some of these subdisciplines are unique to mechanical engineering, while others are a combination of mechanical engineering and one or more other disciplines. Most work that a mechanical engineer does uses skills and techniques from several of these subdisciplines, as well as specialized subdisciplines. Specialized subdisciplines, as used in this article, are usually the subject of graduate studies or on-the-job training more than undergraduate research. Several specialized subdisciplines are discussed at the end of this section.

Mechanics

Mechanics is, in the most general sense, the study of forces and their effect upon matter. Typically, engineering mechanics is used to analyze and predict the acceleration and deformation (both elastic and plastic) of objects under known forces (also called loads) or stresses. Subdisciplines of mechanics include

- Statics, the study of non-moving bodies under known loads
- Dynamics (or kinetics), the study of how forces affect moving bodies
- Mechanics of materials, the study of how different materials deform under various types of stress
- Fluid Mechanics, the study of how fluids react to forces
- Continuum mechanics, a method of applying mechanics that assumes that objects are continuous (rather than discrete)



Mohr's circle, a common tool to study stresses in a mechanical element

Mechanical engineers typically use mechanics in the design or analysis phases of engineering. If the engineering project were the design of a vehicle, statics might be employed to design the frame of the vehicle, in order to evaluate where the stresses will be most intense. Dynamics might be used when designing the car's engine, to evaluate the forces in the pistons and cams as the engine cycles. Mechanics of materials might be used to choose appropriate materials for the frame and engine. Fluid mechanics might be used to design a ventilation system for the vehicle (see HVAC), or to design the intake system for the engine.

Kinematics

Kinematics is the study of the motion of bodies (objects) and systems (groups of objects), while ignoring the forces that cause the motion. The movement of a crane and the oscillations of a piston in an engine are both simple kinematic systems. The crane is a type of open kinematic chain, while the piston is part of a closed four bar linkage.

Mechanical engineers typically use kinematics in the design and analysis of mechanisms. Kinematics can be used to find the possible range of motion for a given mechanism, or, working in reverse, can be used to design a mechanism that has a desired range of motion.

Mechatronics and robotics

Mechatronics is an interdisciplinary branch of mechanical engineering, electrical engineering and software engineering that is concerned with integrating electrical and mechanical engineering to create hybrid systems. In this way, machines can be automated through the use of electric motors, servo-mechanisms, and other electrical systems in conjunction with special software. A common example of a mechatronics system is a CD-ROM drive. Mechanical systems open and close the drive, spin the CD and move the laser, while an optical system reads the data on the CD and converts it to bits. Integrated software controls the process and communicates the contents of the CD to the computer.

Mechatronics is currently used in the following areas of engineering, among others:

- Automation, and in the area of robotics
- Servo-mechanics

- Sensing and control systems
- Automotive engineering, in the design of subsystems such as anti-lock braking systems
- Computer engineering, in the design of mechanisms such as computer drives

Robotics is the application of mechatronics to create robots, which perform tasks that are dangerous, unpleasant, or repetitive. These robots may be of any shape and size, but all are preprogrammed and interact physically with the world. To create a robot, an engineer typically employs kinematics (to determine the robot's range of motion) and mechanics (to determine the stresses within the robot).

Industrial robots perform repetitive tasks, such as assembling vehicles.

Robots are used extensively in Industrial engineering. They allow businesses to save money on labor and perform tasks that are either too dangerous or too precise for humans to perform them economically. Many companies employ assembly lines of robots, and some factories are so robotized that they can run by themselves. Outside the factory, robots have been employed in bomb disposal, space exploration, and many other fields. Robots are also sold for various residential applications.

Structural analysis

Structural analysis is the branch of mechanical engineering (and also civil engineering) devoted to examining why and how objects fail. Structural failures occur in two general modes: static failure, and fatigue failure. *Static structural failure* occurs when, upon being loaded (having a force applied) the object being analyzed either breaks or is deformed plastically, depending on the criterion for failure. *Fatigue failure* occurs when an object fails after a number of repeated loading and unloading cycles. Fatigue failure occurs because of imperfections in the object: a microscopic crack on the surface of the object, for instance, will grow slightly with each cycle (propagation) until the crack is large enough to cause ultimate failure.

Failure is not simply defined as when a part breaks, however; it is defined as when a part does not operate as intended. Some systems, such as the perforated top sections of some plastic bags, are designed to break. If these systems do not break, failure analysis might be employed to determine the cause.

Structural analysis is often used by mechanical engineers after a failure has occurred, or when designing to prevent failure. Engineers may use various books and handbooks such as those published by ASM to aid them in determining the type of failure and possible causes.

Structural analysis may be used in the office when designing parts, in the field to analyze failed parts, or in laboratories where parts might undergo controlled failure tests.

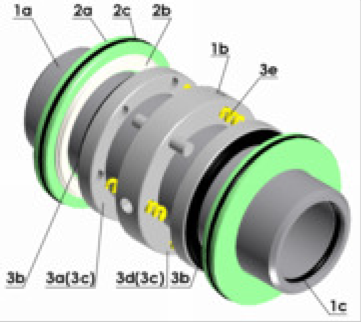
Thermodynamics and thermo-science

Thermodynamics is an applied science used in several branches of engineering, including Mechanical and Chemical Engineering. At its simplest, thermodynamics is the study of energy, its use and transformation through a system. Typically, engineering thermodynamics is concerned with changing energy from one form to another. As an example, automotive engines convert chemical energy (enthalpy) from the fuel into heat, and then into mechanical work that eventually turns the wheels.

Thermodynamics principles are used by mechanical engineers in the fields of heat transfer, thermofluids, and energy conversion. Mechanical engineers use thermo-science to design engines and power plants, heating, ventilation, and air-conditioning (HVAC) systems, heat exchangers, heat sinks, radiators, refrigeration, insulation, and others.

Drafting

Drafting or technical drawing is the means by which mechanical engineers create instructions for manufacturing parts. A technical drawing can be a computer model or hand-drawn schematic showing all the dimensions necessary to manufacture a part, as well as assembly notes, a list of required materials, and other pertinent information. A U.S. mechanical engineer or skilled worker who creates technical drawings may be referred to as a drafter or draftsman (or, in a more politically correct way, draftsperson). Drafting has historically been a two-dimensional process, but recent Computer-Aided Designing (CAD) programs have begun to allow the designer to create in three dimensions.



A CAD model of a mechanical double seal

Instructions for manufacturing a part must be fed to the necessary machinery, either manually, through programmed instructions, or through the use of a Computer-Aided Manufacturing (CAM) or combined CAD/CAM program. Optionally, an engineer may also manually manufacture a part using the technical drawings, but this is becoming an increasing rarity, with the advent of CNC (Computer Numerically Controlled) manufacturing. Engineers primarily manually manufacture parts in the areas of applied spray coatings, finishes, and other processes that cannot economically or practically be done by a machine.

Drafting is used in nearly every subdiscipline of mechanical engineering, and by many other branches of engineering and architecture. Three-dimensional models created using CAD software are also commonly used in Finite element analysis (FEA) and Computational fluid dynamics (CFD).

Specialized subdisciplines

The following is a list of some additional subdisciplines and topics within mechanical engineering. These topics may be considered *specialized* because they are not typically part of undergraduate mechanical engineering requirements, or require training beyond an undergraduate level to be useful.


An aerodynamic test vehicle used by mechanical engineers.

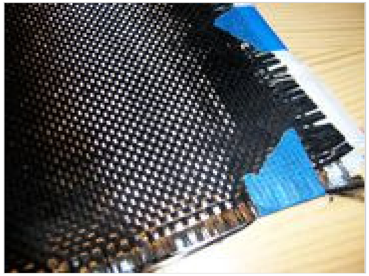
- Acoustical engineering
- Aerospace engineering
- Alternative energy
- Automotive engineering
- Biomedical engineering
- Computer-aided engineering
- Design optimization
- Heating, ventilation, and air conditioning (HVAC)
- Marine engineering
- Nanotechnology
- Nuclear engineering
- Piping

- Power generation
- Engineering-based programming

Frontiers of research

Mechanical engineers are constantly pushing the boundaries of what is physically possible in order to produce safer, cheaper, and more efficient machines and mechanical systems. Some technologies at the cutting edge of mechanical engineering are listed below (see also exploratory engineering).

Composites



Composite cloth consisting of woven carbon fibre.

Composites or composite materials are a combination of materials which provide different physical characteristics than either material separately. Composite material research within mechanical engineering typically focuses on designing (and, subsequently, finding applications for) stronger or more rigid materials while attempting to reduce weight, suceptability to corrosion, and other undesirable factors. Carbon fibre reinforced composites, for instance, have been used in such diverse applications as spacecraft and fishing rods.

Mechatronics

Mechatronics is the synergistic combination of mechanical engineering, electronic engineering, and software engineering. The purpose of this interdisciplinary engineering field is the study of automata from an engineering perspective and serves the purposes of controlling advanced hybrid systems.

Nanotechnology

At the smallest scales, mechanical engineering becomes nanotechnology and molecular engineering—one speculative goal of which is to create a molecular assembler to build molecules and materials via mechanosynthesis. For now this goal remains within exploratory engineering.

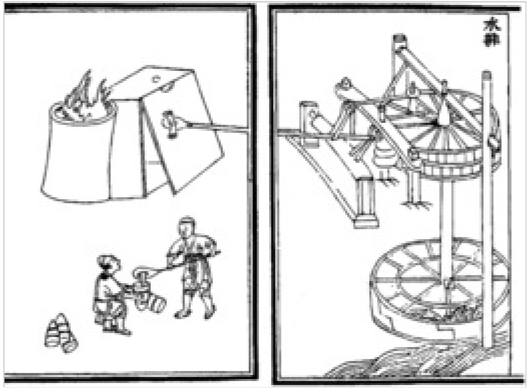
Metallurgy

2008/9 Schools Wikipedia Selection. Related subjects: Engineering; Materials science

Metallurgy is a domain of materials science that studies the physical and chemical behaviour of metallic elements, their intermetallic compounds, and their compounds, which are called alloys. It is also the technology of metals: the way in which science is applied to their practical use. Metallurgy is commonly used in the craft of metalworking.

Georg Agricola, author of *De re metallica*, an important early book on metal extraction

History



An illustration of furnace bellows operated by waterwheels, from the *Nong Shu*, by Wang Zhen, 1313 AD, during the Chinese Yuan Dynasty.

The earliest recorded metal employed by humans appears to be gold which can be found free or "native". Small amounts of natural gold have been found in Spanish caves used during the late Paleolithic period, *c.* 40,000 BC.

Silver, copper, tin and meteoric iron can also be found native, allowing a limited amount of metalworking in early cultures. Egyptian weapons made from meteoric iron in about 3000 B.C. were highly prized as "Daggers from Heaven". However, by learning to get copper and tin by heating rocks and combining copper and tin to make an alloy called bronze, the technology of metallurgy began about 3500 B.C. with the Bronze Age.

Gold headband from Thebes 750-700 BC

The extraction of iron from its ore into a workable metal is much more difficult. It appears to have been invented by the Hittites in about 1200 B.C., beginning the Iron Age. The secret of extracting and working iron was a key factor in the success of the Philistines

Historical developments in ferrous metallurgy can be found in a wide variety of past cultures and civilizations. This includes the ancient and medieval kingdoms and empires of the Middle East and Near East, ancient Egypt and Anatolia (Turkey), Carthage, the Greeks and Romans of ancient Europe, medieval Europe, ancient and medieval China, ancient and medieval India, ancient and medieval Japan, etc. Of interest to note is that many applications, practices, and devices associated or involved in metallurgy were possibly established in ancient China before Europeans mastered these crafts (such as the innovation of the blast furnace, cast iron, steel, hydraulic-powered trip hammers, etc.). However, modern research suggests that Roman technology was far more sophisticated than hitherto supposed, especially in mining methods, metal extraction and forging. They were for example expert in hydraulic mining methods well before the Chinese, or any other civilization of the time.

A 16th century book by Georg Agricola called De re metallica describes the highly developed and complex processes of mining metal ores, metal extraction and metallurgy of the time. Agricola has been described as the "father of metallurgy"

Extractive metallurgy

Extractive metallurgy is the practice of removing valuable metals from an ore and refining the extracted raw metals into a purer form. In order to convert a metal oxide or sulfide to a purer metal, the ore must be reduced either physically, chemically, or electrolytically.

Extractive metallurgists are interested in three primary streams: feed, concentrate (valuable metal oxide/sulfide), and tailings (waste). After mining, large pieces of the ore feed are broken through crushing and/or grinding in order to obtain particles small enough where each particle is either mostly valuable or mostly waste. Concentrating the particles of a value in a form supporting separation enables the desired metal to be removed from waste products.

Mining may not be necessary if the ore body and physical environment are conducive to leaching. Leaching dissolves minerals in an ore body and results in an enriched solution. The solution is collected and processed to extract valuable metals.

Ore bodies often contain more than one valuable metal. Tailings of a previous process may be used as a feed in another process to extract a secondary product from the original ore. Additionally, a concentrate may contain more than one valuable metal. That concentrate would then be processed to separate the valuable metals into individual constituents.

Important common alloy systems

Common engineering metals include aluminium, chromium, copper, iron, magnesium, nickel, titanium and zinc. These are most often used as alloys. Much effort has been placed on understanding the iron-carbon alloy system, which includes steels and cast irons. Plain carbon steels are used in low cost, high strength applications where weight and corrosion are not a problem. Cast irons, including ductile iron are also part of the iron-carbon system.

Stainless steel or galvanized steel are used where resistance to corrosion is important. Aluminium alloys and magnesium alloys are used for applications where strength and lightness are required.

Cupro-nickel alloys such as Monel are used in highly corrosive environments and for non-magnetic applications. Nickel-based superalloys like Inconel are used in high temperature applications such as turbochargers, pressure vessels, and heat exchangers. For extremely high temperatures, single crystal alloys are used to minimize creep.

Production engineering of metals

In production engineering, metallurgy is concerned with the production of metallic components for use in consumer or engineering products. This involves the production of alloys, the shaping, the heat treatment and the surface treatment of the product. The task of the metallurgist is to achieve balance between material properties such as cost, weight, strength, toughness, hardness, corrosion and fatigue resistance, and performance in temperature extremes. To achieve this goal, the operating environment must be carefully considered. In a saltwater environment, ferrous metals and some aluminium alloys corrode quickly. Metals exposed to cold or cryogenic conditions may endure a ductile to brittle transition and lose their toughness, becoming more brittle and prone to cracking. Metals under continual cyclic loading can suffer from metal fatigue. Metals under constant stress at elevated temperatures can creep.

Metal working processes

Metals are shaped by processes such as casting, forging, flow forming, rolling, extrusion, sintering, metalworking, machining and fabrication. With casting, molten metal is poured into a shaped mould. With forging, a red-hot billet is hammered into shape. With rolling, a billet is passed through successively narrower rollers to create a sheet. With extrusion, a hot and malleable metal is forced under pressure through a die, which shapes it before it cools. With sintering, a powdered metal is compressed into a die at high temperature. With machining, lathes, milling machines, and drills cut the cold metal to shape. With fabrication, sheets of metal are cut with guillotines or gas cutters and bent into shape.

" Cold working" processes, where the product’s shape is altered by rolling, fabrication or other processes while the product is cold, can increase the strength of the product by a process called work hardening. Work hardening creates microscopic defects in the metal, which resist further changes of shape.

Various forms of casting exist in industry and academia. These include sand casting, investment casting (also called the “ lost wax process”), die casting and continuous casting.

Joining

Welding

Welding is a technique for joining metal components by melting the base material. A filler material of similar composition may also be melted into the joint.

Brazing

Brazing is a technique for joining metals at a temperature below their melting point. A filler with a melting point below that of the base metal is used, and is drawn into the joint by capillary action. Brazing results in a mechanical and metallurgical bond between work pieces.

Soldering

Soldering is a method of joining metals below their melting points using a filler metal. Soldering results in a mechanical joint and occurs at lower temperatures than brazing.

Heat treatment

Metals can be heat treated to alter the properties of strength, ductility, toughness, hardness or resistance to corrosion. Common heat treatment processes include annealing, precipitation strengthening, quenching, and tempering. The **annealing** process softens the metal by allowing recovery of cold work and grain growth. **Quenching** can be used to harden alloy steels, or in precipitation hardenable alloys, to trap dissolved solute atoms in solution. **Tempering** will cause the dissolved alloying elements to precipitate, or in the case of quenched steels, improve impact strength and ductile properties.

Surface treatment

Plating

Electroplating is a common surface-treatment technique. It involves bonding a thin layer of another metal such as gold, silver, chromium or zinc to the surface of the product. It is used to reduce corrosion as well as to improve the product's aesthetic appearance.

Thermal spray

Thermal spraying techniques are another popular finishing option, and often have better high temperature properties than electroplated coatings.

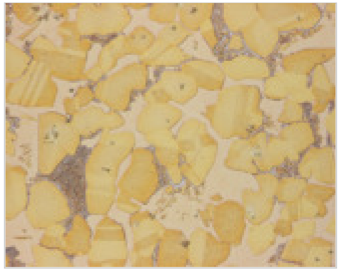
Case hardening

Case hardening is a process in which an alloying element, most commonly carbon or nitrogen, diffuses into the surface of a monolithic metal. The resulting interstitial solid solution is harder than the base material, which improves wear resistance without sacrificing toughness.

Electrical and electronic engineering

Metallurgy is also applied to electrical and electronic materials where metals such as aluminium, copper, tin, silver, and gold are used in power lines, wires, printed circuit boards and integrated circuits.

Metallurgical techniques



Metallography allows the metallurgist to study the microstructure of metals.

Metallurgists study the microscopic and macroscopic properties using metallography, a technique invented by Henry Clifton Sorby. In metallography, an alloy of interest is ground flat and polished to a mirror finish. The sample can then be etched to reveal the microstructure and macrostructure of the metal. A metallurgist can then examine the sample with an optical or electron microscope and learn a great deal about the sample's composition, mechanical properties, and processing history.

Crystallography, often using diffraction of x-rays or electrons, is another valuable tool available to the modern metallurgist. Crystallography allow the identification of unknown materials and reveals the crystal structure of the sample. Quantitative crystallography can be used to calculate the amount of phases present as well as the degree of strain to which a sample has been subjected.

The physical properties of metals can be quantified by mechanical testing. Typical tests include tensile strength, compressive strength, hardness, impact toughness, fatigue and creep life.

Microscope

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Robert Hooke's **microscope** (1665) - an **engineered** device used to study living systems.

A **microscope** (Greek: *μικρόν* (*micron*) = small + *σκοπεῖν* (*skopein*) = to look at) is an instrument for viewing objects that are too small to be seen by the naked or unaided eye. The science of investigating small objects using such an instrument is called microscopy. The term *microscopic* means minute or very small, not visible with the eye unless aided by a microscope. The microscopes used in schools and homes trace their history back almost 400 years.

The first useful microscope was developed in the Netherlands in the early 1600s. There is almost as much confusion about the inventor as about the dates. Three different eyeglass makers have been given credit for the invention: Hans Lippershey (who also developed the first real telescope); Hans Janssen; and his son, Zacharias.

The most common type of microscope—and the first to be invented—is the optical microscope. This is an optical instrument containing one or more lenses that produce an enlarged image of an object placed in the focal plane of the lens(es). There are, however, many other microscope designs.

Types


A 1915 Bausch and Lomb Optical microscope.

Microscopes can largely be separated into two classes, optical theory microscopes and scanning probe microscopes.

Optical theory microscopes are microscopes which function through the optical theory of lenses in order to magnify the image generated by the passage of a wave through the sample. The waves used are either electromagnetic in optical microscopes or electron beams in electron microscopes. The types are the Compound Light, Stereo, and the electron microscope.

Optical microscopes


A stereo microscope is often used for lower-power magnification on large subjects.

Optical microscopes, through their use of visible wavelengths of light, are the simplest and hence most widely used type of microscope. Recent research has shown (see Brian J. Ford's research on simple microscopes) that even simple microscopes, those with a single small lens, gave amazingly clear images to the earliest microscopists. Today **compound microscopes**, i.e., especially those with a series of lenses, serve uses in many fields of science, particularly biology and geology.

Optical microscopes use refractive lenses, typically of glass and occasionally of plastic, to focus light into the eye or another light detector. Typical magnification of a light microscope is up to 1500x with a theoretical resolution of around 0.2 micrometres. Specialised techniques (e.g., scanning confocal microscopy) may exceed this magnification but the resolution is an insurmountable diffraction limit.

Other microscopes which use electromagnetic wavelengths not visible to the human eye are often called optical microscopes. The most common of these, due to its high resolution yet no requirement for a vacuum like electron microscopes, is the x-ray microscope.

Electron microscopes

Electron microscopes, which use beams of electrons instead of light, are designed for very high magnification usage. Electrons, which have a much smaller wavelength than visible light, allow a much higher resolution. The main limitation of the electron beam is that it must pass through a vacuum as air molecules would otherwise scatter the beam.

Instead of relying on refraction, lenses for electron microscopes are specially designed electromagnets which generates magnetic fields that are approximately parallel to the direction that electrons travel. The electrons are typically detected by a phosphor screen, photographic film or a CCD.

Two major variants of electron microscopes exist:

- Scanning electron microscope: looks at the surface of bulk objects by scanning the surface with a fine electron beam and measuring reflection. May also be used for spectroscopy.
- Transmission electron microscope: passes electrons completely through the sample, analogous to basic optical microscopy. This requires careful sample preparation, since electrons are scattered so strongly by most materials. It can also obtain detailed information on the sample's crystallography through selected area diffraction.

Scanning probe microscope

In scanning probe microscopy (SPM), a physical probe is used either in close contact to the sample or nearly touching it. By rastering the probe across the sample, and by measuring the interactions between the sharp tip of the probe and the sample, a micrograph is generated. The exact nature of the interactions between the probe and the sample determines exactly what kind of SPM is being used. Because this kind of microscopy relies on the interactions between the tip and the sample, it generally only measures information about the surface of the sample.

Some kinds of SPMs are:

- Atomic force microscope
- Scanning tunneling microscope
- Electric force microscope
- Magnetic force microscope (MFM)
- Near-field scanning optical microscope

Point-projection microscopes

The field emission microscope, field ion microscope, and the Atom Probe are examples of point-projection microscopes where ions are excited from a needle-shaped specimen and hit a detector. The Atom-Probe Tomograph (APT) is the most modern incarnation and allows a three-dimensional atom-by-atom (with chemical elements identified) reconstruction with sub-nanometer resolution.

Other microscopes

Acoustic microscopes use sound waves to measure variations in acoustic impedance. Similar to SONAR in principle, they are used for such jobs as detecting defects in the subsurfaces of materials including those found in integrated circuits.

Milky Way

2008/9 Schools Wikipedia Selection. Related subjects: Space (Astronomy)

The **Milky Way** (a translation of the Latin *Via Lactea*, in turn derived from the Greek *Γαλαξίας* (Galaxias) sometimes referred to simply as " the Galaxy"), is a barred spiral galaxy that is part of the Local Group of galaxies. Although the Milky Way is one of billions of galaxies in the observable universe, the Galaxy has special significance to humanity as it is the home galaxy of the Solar System. The plane of the Milky Way galaxy is visible from Earth as a band of light in the night sky, and it is the appearance of this band of light which has inspired the name for our galaxy.

Some sources hold that, strictly speaking, the term *Milky Way* should refer exclusively to the observation of the band of light, while the full name *Milky Way Galaxy*, or alternatively *the Galaxy* should be used to describe our galaxy as an astrophysical whole. It is unclear how widespread the usage of this convention is, however, and the term *Milky Way* is routinely used in either context.

Milky Way Galaxy

Infrared image of the core of the Milky Way galaxy

Observation data	
Type	SBbc (barred spiral galaxy)
Diameter	100,000 light years
Thickness	12,000 light years (gas) 1,000 light years (stars)
Number of stars	200 to 400 billion
Oldest star	13.2 billion years
Mass	5.8×10^{11} $M_\odot$
Sun's distance to galactic centre	26,000 ± 1400 light-years
Sun's galactic rotation period	220 million years (negative rotation)
Spiral pattern rotation period	50 million years

View from Earth

The Milky Way galaxy, as viewed from the Earth, itself situated on one of the spiral arms of the galaxy (see Sun's location), appears as a hazy band of white light in the night sky arching across the entire celestial sphere originating from stars and other material which lie within the galactic plane. The plane of the Milky Way is inclined by about 60° to the ecliptic (the plane of the earth's orbit), with the North Galactic Pole situated at right ascension 12h49m, declination +27.4° (B1950) near beta Comae Berenices. The South Galactic Pole is near alpha Sculptoris.

The centre of the galaxy is in the direction of Sagittarius, and the Milky Way then "passes" (going westward) through Scorpius, Ara, Norma, Triangulum Australe, Circinus, Centaurus, Musca, Crux, Carina, Vela, Puppis, Canis Major, Monoceros, Orion & Gemini, Taurus, Auriga, Perseus, Andromeda, Cassiopeia, Cepheus & Lacerta, Cygnus, Vulpecula, Sagitta, Aquila, Ophiuchus, Scutum, and back to Sagittarius.

The Milky Way looks brightest in the direction of the constellation of Sagittarius, toward the galactic centre. Relative to the celestial equator, it passes as far north as the constellation of Cassiopeia and as far south as the constellation of Crux, indicating the high inclination of Earth's equatorial plane and the plane of the ecliptic relative to the galactic plane. The fact that the Milky Way divides the night sky into two roughly equal hemispheres indicates that our Solar System lies close to the galactic plane. The Milky Way has a relatively low surface brightness, making it difficult to see from any urban or suburban location suffering from light pollution.

Size

Bar pattern rotation period	15 to 18 million years
Speed relative to CMB rest frame	552 km/s

See also: Galaxy, List of galaxies

The stellar disk of the Milky Way galaxy is approximately 100,000 light years in diameter, and is believed to be, on average, about 1,000 light years thick. It is estimated to contain at least 200 billion stars and possibly up to 400 billion stars, the exact figure depending on the number of very low-mass stars, which is highly uncertain. Extending beyond the stellar disk is a much thicker disk of gas. Recent observations indicate that the gaseous disk of the Milky Way has a thickness of around 12,000 light years - twice the previously accepted value. As a guide to the relative physical scale of the Milky Way, if it were reduced to 130 km (80 mi) in diameter, the Solar System would be a mere 2 mm (0.08 inches) in width.

The Galactic Halo extends outward, but is limited in size by the orbits of the two Milky Way satellites, the Large and the Small Magellanic Clouds, whose perigalacticon is at ~180,000 light-years.

Age

It is extremely difficult to define the age at which the Milky Way formed, but the age of the oldest star in the Galaxy yet discovered is estimated to be about 13.2 billion years, nearly as old as the Universe itself.

A green and red Perseid meteor is striking the sky just below the Milky Way in August 2007.

This estimate is based on research by a team of astronomers in 2004 using the UV-Visual Echelle Spectrograph of the Very Large Telescope to measure, for the first time, the beryllium content of two stars in globular cluster NGC 6397. From this research, the elapsed time between the rise of the first generation of stars in the entire Galaxy and the first generation of stars in the cluster was deduced to be 200 million to 300 million years. By including the estimated age of the stars in the globular cluster (13.4 ± 0.8 billion years), they estimated the age of the oldest stars in the Milky Way at 13.6 ± 0.8 billion years. Based upon this emerging science, the Galactic thin disk is estimated to have been formed between 6.5 and 10.1 billion years ago.

Composition and structure

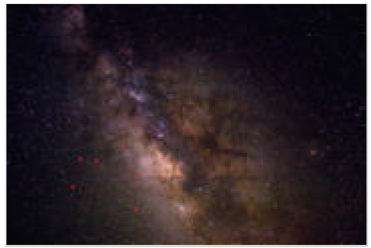
The Galaxy consists of a bar-shaped core region surrounded by a disk of gas, dust and stars forming four distinct arm structures spiralling outward in a logarithmic spiral shape (see Spiral arms). The mass distribution within the Galaxy closely resembles the Sbc Hubble classification, which is a spiral galaxy with relatively loosely-wound arms. Astronomers in the 1980s first began to suspect that the Milky Way is a barred spiral galaxy rather than an ordinary spiral galaxy and their suspicions were confirmed by the Spitzer Space Telescope observations in 2005 which showed the Galaxy's central bar to be larger than previously suspected. As of 2006, the Milky Way's mass is thought to be about 5.8×10^{11} solar masses ($M_\odot$) comprising 200 to 400 billion stars. Its integrated absolute visual magnitude has been estimated to be −20.9. Most of the mass of the Galaxy is thought to be dark matter, forming a dark matter halo of an estimated 600–3000 billion $M_\odot$ which is spread out relatively evenly.

Studies in 2008 have suggested that the Milky Way is a barred spiral galaxy. The authors suggest Messier 109 as one possible analog.

Galactic centre

The galactic disc, which bulges outward at the galactic centre, has a diameter of between 70,000 and 100,000 light-years. The distance from the Sun to the galactic centre is now estimated at 26,000 ± 1400 light-years, while older estimates could put the Sun as far as 35,000 light-years from the central bulge.

The galactic centre harbors a compact object of very large mass (named Sagittarius A*), strongly suspected to be a supermassive black hole. Most galaxies are believed to have a supermassive black hole at their centre.

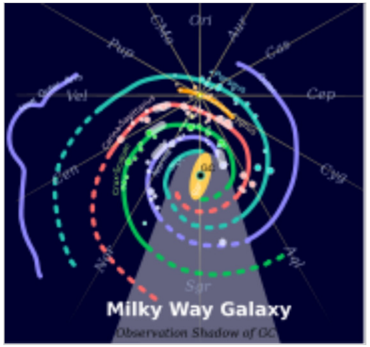


The galactic centre in the direction of Sagittarius. The primary stars of Sagittarius are indicated in red.

The Galaxy's bar is thought to be about 27,000 light-years long, running through its center at a 44 ± 10 degree angle to the line between the Sun and the centre of the Galaxy. It is composed primarily of red stars, believed to be ancient (see red dwarf, red giant). The bar is surrounded by a ring called the "5- kpc ring" that contains a large fraction of the molecular hydrogen present in the Galaxy, as well as most of the Milky Way's star formation activity. Viewed from the Andromeda Galaxy, it would be the brightest feature of our own galaxy.

Spiral arms

Each spiral arm describes a logarithmic spiral (as do the arms of all spiral galaxies) with a pitch of approximately 12 degrees.
There are believed to be four major spiral arms which all start at the Galaxy's centre. These are named as follows, according to the image at left:



Observed and extrapolated structure of the spiral arms

colour	arm(s)
cyan	3-k pc and Perseus Arm
purple	Norma and Cygnus Arm (Along with a newly discovered extension)
green	Scutum-Crux Arm
pink	Carina and Sagittarius Arm
There are at least two smaller arms or spurs, including:	
orange	Orion Arm (which contains our own Solar System and Sun)

Outside of the major spiral arms is the Outer Ring or Monoceros Ring, a ring of stars around the Milky Way proposed by astronomers Brian Yanny and Heidi Jo Newberg, which consists of gas and stars torn from other galaxies billions of years ago.

As is typical for many galaxies, the distribution of mass in the Milky Way Galaxy is such that the orbital speed of most stars in the Galaxy does not depend strongly on its distance from the centre. Away from the central bulge or outer rim, the typical stellar velocity is between 210 and 240 km/s. Hence the orbital period of the typical star is directly proportional only to the length of the path traveled. This is unlike the situation within the Solar System, where two-body gravitational dynamics dominate and different orbits are expected to have significantly different velocities associated with them. This difference is one of the major pieces of evidence for the existence of dark matter. Another interesting aspect is the so-called "wind-up problem" of the spiral arms. If one believes that the inner parts of the arms rotate faster than the outer part, then the Galaxy will wind up so much that the spiral structure will be thinned out. But this is not what is observed in spiral galaxies; instead, astronomers propose that the spiral arms form as a result of a matter-density wave emanating from the galactic centre. This can be likened to a moving traffic jam on a highway — the cars are all moving, but there is always a region of slow-moving cars. Thus this results in several spiral arms where there are a lot of stars and gas. This model also agrees with enhanced star formation in or near spiral arms; the compressional waves increase the density of molecular hydrogen and protostars form as a result.

Observations presented in 2008 by Robert Benjamin of the University of Wisconsin-Whitewater suggest that the Milky Way possesses only two major stellar arms: the Perseus arm and the Scutum-Centaurus arm. The rest of the arms are minor or adjunct arms.

Artist's conception of the spiral structure of the Milky Way with two major, stellar arms and a bar.

Halo

The galactic disk is surrounded by a spheroid halo of old stars and globular clusters, of which 90% lie within 100,000 light-years, suggesting a stellar halo diameter of 200,000 light-years. However, a few globular clusters have been found farther, such as PAL 4 and AM1 at more than 200,000 light-years away from the galactic centre. While the disk contains gas and dust which obscure the view in some wavelengths, the spheroid component does not. Active star formation takes place in the disk (especially in the spiral arms, which represent areas of high density), but not in the halo. Open clusters also occur primarily in the disk.

Recent discoveries have added dimension to the knowledge of the Milky Way's structure. With the discovery that the disc of the Andromeda Galaxy (M31) extends much further than previously thought, the possibility of the disk of our own Galaxy extending further is apparent, and this is supported by evidence of the newly discovered Outer Arm extension of the Cygnus Arm. With the discovery of the Sagittarius Dwarf Elliptical Galaxy came the discovery of a ribbon of galactic debris as the polar orbit of Sagittarius and its interaction with the Milky Way tears it apart. Similarly, with the discovery of the Canis Major Dwarf Galaxy, it was found that a ring of galactic debris from its interaction with the Milky Way encircles the galactic disk.

On January 9, 2006, Mario Juric and others of Princeton University announced that the Sloan Digital Sky Survey of the northern sky found a huge and diffuse structure (spread out across an area around 5,000 times the size of a full moon) within the Milky Way that does not seem to fit within current models. The collection of stars rises close to perpendicular to the plane of the spiral arms of the Galaxy. The proposed likely interpretation is that a dwarf galaxy is merging with the Milky Way. This galaxy is tentatively named the Virgo Stellar Stream and is found in the direction of Virgo about 30,000 light-years away.

Sun's location

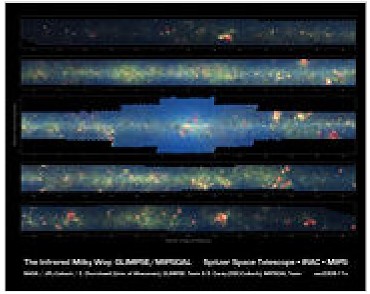
The Sun (and therefore the Earth and Solar System) may be found close to the inner rim of the Galaxy's Orion Arm, in the Local Fluff or the Gould Belt, at a hypothesized distance of 7.62±0.32 kpc from the Galactic Centre. The distance between the local arm and the next arm out, the Perseus Arm, is about 6,500 light-years. The Sun, and thus the Solar System, is found in what scientists call the galactic habitable zone.

The Apex of the Sun's Way, or the solar apex, is the direction that the Sun travels through space in the Milky Way. The general direction of the Sun's galactic motion is towards the star Vega near the constellation of Hercules, at an angle of roughly 60 sky degrees to the direction of the Galactic Centre. The Sun's orbit around the Galaxy is expected to be roughly elliptical with the addition of perturbations due to the galactic spiral arms and non-uniform mass distributions. In addition the Sun oscillates up and down relative to the galactic plane approximately 2.7 times per orbit. This is very similar to how a simple harmonic oscillator works with no drag force (damping) term. Due to the higher density of stars close to the galactic plane, these oscillations often coincide with mass extinction periods on Earth, presumably due to increased impact events.

It takes the Solar System about 225–250 million years to complete one orbit of the galaxy (a *galactic year*), so it is thought to have completed 20–25 orbits during the lifetime of the Sun and 1/1250th of a revolution since the origin of humans. The orbital speed of the Solar System about the centre of the Galaxy is approximately 220 km/s. At this speed, it takes around 1400 years for the Solar System to travel a distance of 1 light-year, or 8 days to travel 1 AU.

Environment

The Milky Way and the Andromeda Galaxy are a binary system of giant spiral galaxies belonging to a group of 50 closely bound galaxies known as the Local Group, itself being part of the Virgo Supercluster.



Broad infrared view of our Milky Way Galaxy from the Spitzer Space Telescope created from more than 800,000 frames. This is the most detailed infrared picture of our galaxy to date.

Two smaller galaxies and a number of dwarf galaxies in the Local Group orbit the Milky Way. The largest of these is the Large Magellanic Cloud with a diameter of 20,000 light-years. It has a close companion, the Small Magellanic Cloud. The Magellanic Stream is a peculiar streamer of neutral hydrogen gas connecting these two small galaxies. The stream is thought to have been dragged from the Magellanic Clouds in tidal interactions with the Galaxy. Some of the dwarf galaxies orbiting the Milky Way are Canis Major Dwarf (the closest), Sagittarius Dwarf Elliptical Galaxy, Ursa Minor Dwarf, Sculptor Dwarf, Sextans Dwarf, Fornax Dwarf, and Leo I Dwarf. The smallest Milky Way dwarf galaxies are only 500 light-years in diameter. These include Carina Dwarf, Draco Dwarf, and Leo II Dwarf. There may still be undetected dwarf galaxies, which are dynamically bound to the Milky Way. Observations through the zone of avoidance are frequently detecting new distant and nearby galaxies. Some galaxies consisting mostly of gas and dust may also have evaded detection so far.

In January 2006, researchers reported that the heretofore unexplained warp in the disk of the Milky Way has now been mapped and found to be a ripple or vibration set up by the Large and Small Magellanic Clouds as they circle the Galaxy, causing vibrations at certain frequencies when they pass through its edges. Previously, these two galaxies, at around 2% of the mass of the Milky Way, were considered too small to influence the Milky Way. However, by taking into account dark matter, the movement of these two galaxies creates a wake that influences the larger Milky Way. Taking dark matter into account results in an approximately twenty-fold increase in mass for the Galaxy. This calculation is according to a computer model made by Martin Weinberg of the University of Massachusetts, Amherst. In this model, the dark matter is spreading out from the galactic disc with the known gas layer. As a result, the model predicts that the gravitational effect of the Magellanic Clouds is amplified as they pass through the Galaxy.

Current measurements suggest the Andromeda Galaxy is approaching us at 100 to 140 kilometers per second. The Milky Way may collide with it in 3 to 4 billion years, depending on the importance of unknown lateral components to the galaxies' relative motion. If they collide, it is thought that the Sun and the other stars in the Milky Way will probably not collide with the stars of the Andromeda Galaxy, but that the two galaxies will merge to form a single elliptical galaxy over the course of about a billion years.

Velocity

In the general sense, the absolute velocity of any object through space is not a meaningful question according to Einstein's special theory of relativity, which declares that there is no "preferred" inertial frame of reference in space with which to compare the Galaxy's motion. (Motion must always be specified with respect to another object.)

Astronomers believe the Milky Way is moving at approximately 600 km per second relative to the observed locations of other nearby galaxies. Most recent estimates range from 130 km/s to 1,000 km/s. If the Galaxy is moving at 600 km/s, Earth travels 51.84 million km per day, or more than 18.9 billion km per year, about 4.5 times its closest distance from Pluto. The Galaxy is thought to be moving towards the constellation Hydra, and may someday become a close-knit member of the Virgo cluster of galaxies.

Another reference frame is provided by the cosmic microwave background (CMB). The Milky Way appears to be moving at around 552 km/s with respect to the photons of the CMB. This motion is observed by satellites such as COBE and WMAP as a dipole contribution to the CMB, as photons in equilibrium in the CMB frame get blue-shifted in the direction of the motion and red-shifted in the opposite direction.

History

Etymology and beliefs

There are many creation myths around the world which explain the origin of the Milky Way and give it its name. The English phrase is a translation from Greek Γαλαξίας, *Galaxias*, which is derived from the word for milk (γάλα, *gala*). This is also the origin of the word *galaxy*. In Greek myth, the Milky Way was caused by milk spilt by Hera when suckling Heracles.

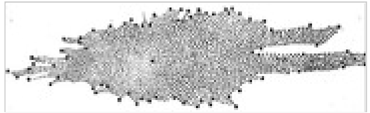

Jacopo Tintoretto's "*The Origin of the Milky Way*"

The term *Milky Way* first appeared in English literature in a poem by Chaucer.

> "See yonder, lo, the Galaxyë
> Which men clepeth the Milky Wey,
> For hit is whyt."
>
> —Geoffrey Chaucer, Geoffrey Chaucer *The House of Fame*, *c.* 1380.

In a large area from Central Asia to Africa, the name for the Milky Way is related to the word for straw. It has been claimed that this was spread by Arabs who in turn borrowed the word from Armenian. In several Uralic, Turkic languages, Fenno-Ugric languages and in the Baltic languages the Milky Way is called the "Birds' Path". The Chinese name "Silver River" (銀河) is used throughout East Asia, including Korea and Japan. An alternative name for the Milky Way in ancient China, especially in poems, is "Heavenly River of Han"(天汉). In Japanese, "Silver River" (銀河 *ginga*) means galaxies in general and the Milky Way is called the "Silver River System" (銀河系 *gingakei*) or the "River of Heaven" (天の川 *ama no* *kawa*). In Swedish, it is called *Vintergatan*, or "Winter Street", because the stars in the belt were used to predict time of the approaching winter.

Discovery



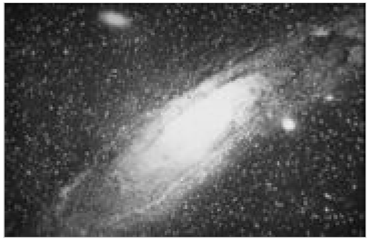
The shape of the Milky Way as deduced from star counts by William Herschel in 1785; the Solar System was assumed near centre.

As Aristotle (384-322 BC) informs us in *Meteorologica* (DK 59 A80), the Greek philosophers Anaxagoras (ca. 500–428 BC) and Democritus (450–370 BC) proposed that the Milky Way might consist of distant stars. However, Aristotle himself believed the Milky Way to be caused by "the ignition of the fiery exhalation of some stars which were large, numerous and close together" and that the "ignition takes place in the upper part of the atmosphere, in the region of the world which is continuous with the heavenly motions." The Arabian astronomer, Alhazen (965-1037 AD), refuted this by making the first attempt at observing and measuring the Milky Way's parallax, and he thus "determined that because the Milky Way had no parallax, it was very remote from the earth and did not belong to the atmosphere."

The Persian astronomer, Abū Rayhān al-Bīrūnī (973-1048), proposed the Milky Way galaxy to be a collection of countless nebulous stars. Avempace (d. 1138) proposed the Milky Way to be made up of many stars but appears to be a continuous image due to the effect of refraction in the Earth's atmosphere. Ibn Qayyim Al-Jawziyya (1292-1350) proposed the Milky Way galaxy to be "a myriad of tiny stars packed together in the sphere of the fixed stars" and that that these stars are larger than planets.

Actual proof of the Milky Way consisting of many stars came in 1610 when Galileo Galilei used a telescope to study the Milky Way and discovered that it was composed of a huge number of faint stars. In a treatise in 1755, Immanuel Kant, drawing on earlier work by Thomas Wright, speculated (correctly) that the Milky Way might be a rotating body of a huge number of stars, held together by gravitational forces akin to the Solar System but on much larger scales. The resulting disk of stars would be seen as a band on the sky from our perspective inside the disk. Kant also conjectured that some of the nebulae visible in the night sky might be separate "galaxies" themselves, similar to our own.

The first attempt to describe the shape of the Milky Way and the position of the Sun within it was carried out by William Herschel in 1785 by carefully counting the number of stars in different regions of the sky. He produced a diagram of the shape of the Galaxy with the Solar System close to the centre.


Photograph of the "Great Andromeda Nebula" from 1899, later identified as the Andromeda Galaxy

In 1845, Lord Rosse constructed a new telescope and was able to distinguish between elliptical and spiral-shaped nebulae. He also managed to make out individual point sources in some of these nebulae, lending credence to Kant's earlier conjecture.

In 1917, Heber Curtis had observed the nova S Andromedae within the "Great Andromeda Nebula" (Messier object M31). Searching the photographic record, he found 11 more novae. Curtis noticed that these novae were, on average, 10 magnitudes fainter than those that occurred within our galaxy. As a result he was able to come up with a distance estimate of 150,000 parsecs. He became a proponent of the "island universes" hypothesis, which held that the spiral nebulae were actually independent galaxies. In 1920 the Great Debate took place between Harlow Shapley and Heber Curtis, concerning the nature of the Milky Way, spiral nebulae, and the dimensions of the universe. To support his claim that the Great Andromeda Nebula was an external galaxy, Curtis noted the appearance of dark lanes resembling the dust clouds in the Milky Way, as well as the significant Doppler shift.

The matter was conclusively settled by Edwin Hubble in the early 1920s using a new telescope. He was able to resolve the outer parts of some spiral nebulae as collections of individual stars and identified some Cepheid variables, thus allowing him to estimate the distance to the nebulae: they were far too distant to be part of the Milky Way. In 1936 Hubble produced a classification system for galaxies that is used to this day, the Hubble sequence.

Mobile phone

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

The **mobile phone** or **mobile**, also called a **cellular phone**, or **cell phone** is a long-range, portable electronic device used for mobile communication that uses a network of specialized base stations known as cell sites. In addition to the standard voice function of a telephone, current mobile phones can support many additional services such as SMS for text messaging, email, packet switching for access to the Internet, and MMS for sending and receiving photos and video. Most current mobile phones connect to a cellular network of base stations (cell sites), which is in turn interconnected to the public switched telephone network (PSTN) (the exception is satellite phones).

Mobile communication standards

GSM / UMTS (3GPP) Family

GSM (2G)

- GPRS
- EDGE (EGPRS)
  - EDGE Evolution
- CSD
  - HSCSD

UMTS (3G)

- HSPA
  - HSDPA
  - HSUPA
  - HSPA+
- UMTS-TDD
  - TD-CDMA
  - TD-SCDMA
- FOMA

UMTS Rev. 8 (Pre-4G)

- LTE
- HSOPA (Super 3G)

cdmaOne / CDMA2000 (3GPP2) Family

cdmaOne (2G)

CDMA2000 (3G)

- EV-DO

UMB (Pre-4G)

AMPS Family

AMPS (1G)

- TACS / ETACS

D-AMPS (2G)

Other Technologies

0G

- PTT
- MTS
- IMTS
- AMTS
- OLT
- MTD
- Autotel / PALM
- ARP

1G

- NMT
- Hicap
- CDPD
- Mobitex
- DataTAC

2G

- iDEN
- PDC

History

- CSD
- PHS
- WiDEN

Pre-4G

- iBurst
- HIPERMAN
- WiMAX
- WiBro
- GAN (UMA)

Channel Access Methods

- FDMA
  - OFDMA
- TDMA
- SSMA
  - CDMA

Frequency bands

- Cellular
  - GSM
  - UMTS
  - PCS
- SMR

Several examples of non-flip mobile phones.

Various cell phones from 1992 to 2004.
Legend:
1. NEC Cellstar 500 series (1992)
2. Nokia 2110 series (1994)
3. Nokia 5120 (1998)
4. Kyocera 2135 (2002)
5. Audiovox CDM8300 (2002)
6. Samsung SCH-A650 (2004)

There is one U.S. patent, Patent Number 887357 for a wireless telephone, issued 1908 to Nathan B. Stubblefield of Murray, Kentucky. He applied this to "cave radio" telephones and not directly to cellular telephony as the term is currently understood. However, the introduction of cells for mobile phone base stations, invented in 1947 by Bell Labs engineers at AT&T, was further developed by Bell Labs during the 1960s. Radiophones have a long and varied history going back to Reginald Fessenden's invention and shore-to-ship demonstration of radio telephony, through the Second World War with military use of radio telephony links and civil services in the 1950s, while hand-held cellular radio devices have been available since 1973. Due to their low establishment costs and rapid deployment, mobile phone networks have since spread rapidly throughout the world, outstripping the growth of fixed telephony.

In 1945, the zero generation (0G) of mobile telephones was introduced. 0G mobile telephones, such as Mobile Telephone Service, were not officially categorized as mobile phones, since they did not support the automatic change of channel frequency during calls, which allows the user to move from one cell (the base station coverage area) to another cell, a feature called " handover".

In 1984, Bell Labs invented such a "call handoff" feature, which allowed mobile-phone users to travel through several cells during the same conversation. Motorola is widely considered to be the inventor of the first practical mobile phone for handheld use in a non-vehicle setting. Using a modern, if somewhat heavy portable handset, Motorola manager Martin Cooper made the first call on a handheld mobile phone on April 3, 1973.

The first commercial cellular network was launched in Japan by NTT in 1979. Fully automatic cellular networks were first introduced in the early to mid 1980s (the 1G generation) with the Nordic Mobile Telephone (NMT) system in 1981. This was followed by a boom in mobile telephone usage, particularly in Northern Europe.

The first "modern" network technology on digital 2G (second generation) cellular technology was launched by Radiolinja (now part of Elisa Group) in 1991 in Finland on the GSM standard which also marked the introduction of competition in mobile telecoms when Radiolinja challenged incumbent Telecom Finland (now part of TeliaSonera) who ran a 1G NMT network. A decade later, the first commercial launch of 3G (Third Generation) was again in Japan by NTT DoCoMo on the WCDMA standard. Until the early 1990s, most mobile phones were too large to be carried in a jacket pocket, so they were typically installed in vehicles as car phones. With the miniaturization of digital components, mobile phones have become increasingly portable over the years.

Today, video and TV services are driving forward third generation (3G) deployment. And in the future, low cost, high speed data will drive forward the fourth generation (4G) as short-range communication emerges. Service and application ubiquity, with a high degree of personalization and synchronization between various user appliances,will be another driver. At the same time, it is probable that the radio access network will evolve from a centralized architecture to a distributed one.

Manufacturers

Nokia Corporation is currently the world's largest manufacturer of mobile telephones, with a global device market share of approximately 36% in Q1 of 2007. Other mobile phone manufacturers include Apple Inc., Audiovox (now UTStarcom), Benefon, BenQ-Siemens, High Tech Computer Corporation (HTC), Fujitsu, Kyocera, LG Mobile, Mitsubishi, Motorola, NEC, Neonode, Panasonic (Matsushita Electric), Pantech Curitel, Philips, Research In Motion, Sagem, Samsung, Sanyo, Sharp, Siemens, Sierra Wireless, SK Teletech, Sonim Technologies, Sony Ericsson, T&A Alcatel, and Toshiba. There are also specialist communication systems related to (but distinct from) mobile phones.

The mobile phone manufacturers can be grouped into two. The top five are available in practically all countries and comprise about 75% of all phones sold. A second tier of small manufacturers exists with phones mostly sold only in specific regions or for niche markets. The top five in order of market share are Nokia, Samsung, Motorola, SonyEricsson and LG.

Subscriptions


This Railfone found on some Amtrak trains uses cellular technology.

Several countries, including the UK, now have more mobile phones than people. There are over five hundred million active mobile phone accounts in China, as of 2007. Luxembourg has the highest mobile phone penetration rate in the world, at 164% in December 2001. In Hong Kong the penetration rate reached 139.8% of the population in July 2007. The total number of mobile phone subscribers in the world was estimated at 2.14 billion in 2005. The subscriber count reached 2.7 billion by end of 2006 according to Informa, and 3.3 billion by November, 2007, thus reaching an equivalent of over half the planet's population. Around 80% of the world's population enjoys mobile phone coverage as of 2006. This figure is expected to increase to 90% by the year 2010.

At present, Africa has the largest growth rate of cellular subscribers in the world, its markets expanding nearly twice as fast as Asian markets. The availability of prepaid or 'pay-as-you-go' services, where the subscriber is not committed to a long term contract, has helped fuel this growth in Africa as well as in other continents.

On a numerical basis, India is the largest growth market, adding about 6 million cell phones every month. With 256.55 million cell phones, market penetration in the country is still low at 22.52%. India expects to reach 500 million subscribers by end of 2010.

There are three major technical standards for the current generation of mobile phones and networks, and two major standards for the next generation 3G phones and networks. All European, African and many Asian countries have adopted a single system, GSM, which is the only technology available on all continents and in most countries and covers over 74% of all subscribers on mobile networks. In many countries, such as the United States, Australia, Brazil, India, Japan, and South Korea GSM co-exists with other internationally adopted standards such as CDMA and TDMA, as well as national standards such as iDEN in the USA and PDC in Japan. Over the past five years several dozen mobile operators (carriers) have abandoned networks on TDMA and CDMA technologies, switching over to GSM.

With third generation (3G) networks, which are also known as IMT-2000 networks, about three out of four networks are on the W-CDMA (also known as UMTS) standard, usually seen as the natural evolution path for GSM and TDMA networks. One in four 3G networks is on the CDMA2000 1x EV-DO technology. Some analysts count a previous stage in CDMA evolution, CDMA2000 1x RTT, as a 3G technology whereas most standardization experts count only CDMA2000 1x EV-DO as a true 3G technology. Because of this difference in interpreting what is 3G, there is a wide variety in subscriber counts. As of June 2007, on the narrow definition there are 200 million subscribers on 3G networks. By using the more broad definition, the total subscriber count of 3G phone users is 475 million.

While some systems of payment are 'pay-as-you-go' where conversation time is purchased and added to a phone unit via an Internet account or in shops or ATMs, other systems are more traditional ones where bills are paid by regular intervals. Pay as you go (also known as "pre-pay") accounts were invented simultaneously in Portugal and Italy and today form more than half of all mobile phone subscriptions. USA, Canada, Japan and Finland are among the rare countries left where most phones are still contract-based.

Culture and customs

In less than twenty years, the mobile telephone has gone from being rare, expensive equipment of the business elite to a pervasive, low-cost personal item. In many countries, mobile telephones outnumber land-line telephones; in the U.S., 50 percent of children have mobile telephones. In many young adults' households it has supplanted the land-line telephone. The mobile phone is banned in some countries, such as North Korea.

Given the high levels of societal mobile telephone service penetration, it is a key means for people to communicate with each other. The SMS feature spawned the " texting" sub-culture. In December 1993, the first person-to-person SMS text message was transmitted in Finland. Currently, texting is the most widely-used data service; 1.8 billion users generated $80 billion of revenue in 2006 (source ITU).

Many telephones offer Instant Messenger services for simple, easy texting. Mobile phones have Internet service (e.g. NTT DoCoMo's i-mode), offering text messaging via e-mail in Japan, South Korea, China, and India. In Europe, 30–40 per cent of internet access is via mobile telephone. Most mobile internet access is much different from computer access, featuring alerts, weather data, e-mail, search engines, instant messages, and game and music downloading; most mobile internet access is hurried and short.

The mobile telephone can be a fashion totem custom-decorated to reflect the owner's personality. This aspect of the mobile telephony business is, in itself, an industry, e.g. ringtone sales exceeded $5 billion in 2006.

Etiquette

The use of a mobile phone is prohibited in some train company carriages

Mobile telephone use in etiquette is an important matter of social discourtesy, phones ringing during funerals, weddings, in toilets, cinemas, and plays. Users often speak loudly, leading to book shops, libraries, bathrooms, cinemas, doctors' offices, and places of worship prohibiting their uses, and, in some places, the installation of signal-jamming equipment to prevent their use (though in many countries, including the U.S., such equipment is currently illegal). Some new buildings, such as auditoriums, have installed wire mesh in the walls (making it a Faraday cage), which prevents signal penetration without violating signal jamming laws.

Trains, particularly those involving long-distance services, often offer a "quiet carriage" where phone use is prohibited, much like the designated non-smoking carriage in the past. However many users tend to ignore this as it is rarely enforced, especially if the other carriages are crowded and they have no choice but to go in the "quiet carriage". Mobile phone use on aircraft is also prohibited and many airlines claim in their in-plane announcements that this prohibition is due to possible interference with aircraft radio communications. Shut-off mobile phones do not interfere with aircraft avionics; the concern is partially based on the crash of Crossair Flight 498.

As of 2007, several airlines are experimenting with base station and antenna systems installed to the aeroplane, allowing low power, short-range connection of any phones aboard to remain connected to the aircraft's base station. Thus, they would not attempt connection to the ground base stations as during take off and landing. Simultaneously, airlines may offer phone services to their travelling passengers either as full voice and data services, or initially only as SMS text messaging and similar services. Qantas, the Australian airline, is the first airline to run a test aeroplane in this configuration in the autumn of 2007. Emirates has announced plans to allow limited mobile phone usage on some flights.

In any case, there are inconsistencies between practices allowed by different airlines and even on the same airline in different countries. For example, Northwest Airlines may allow the use of mobile phones immediately after landing on a domestic flight within the US, whereas they may state "not until the doors are open" on an international flight arriving in the Netherlands. In April 2007 the US Federal Communications Commission officially grounded the idea of allowing passengers to use phones during a flight.

In a similar vein, signs are put up in many countries, such as Canada, the U.K. and the U.S., at petrol stations prohibiting the use of mobile phones, due to possible safety issues. Most schools in the United States have prohibited mobile phones in the classroom, due to the large number of class disruptions that result from their use, the potential for cheating via text messaging, and the possibility of photographing someone without consent. In the UK, possession of a mobile phone in an examination can result in immediate disqualification from that subject or from all that student's subjects.

A working group, made up of Finnish telephone companies, public transport operators and communications authorities, have launched a campaign to remind mobile phone users of courtesy, especially when using mass transit – what to talk about on the phone, and how to. In particular, the campaign wants to impact loud mobile phone usage as well as calls regarding sensitive matters.

Many US cities with subway transit systems underground are studying or have implemented cell phone reception in their underground tunnels for their riders. Boston, Massachusetts has investigated such usage in their tunnels, although there is a question of usage etiquette and also how to fairly award contracts to carriers.

The issue of mobile communication and etiquette has also become an issue of academic interest. The rapid adoption of the device has resulted in the intrusion of telephony into situations where this was previously not known. This has exposed the implicit rules of courtesy and opened them to reevaluation. http://www.richardling.com/papers/1997_One_can_talk_about_common_manners.pdf

Use in disaster response

The Finnish government decided in 2005 that the fastest way to warn citizens of disasters was the mobile phone network. In Japan, mobile phone companies provide immediate notification of earthquakes and other natural disasters to their customers free of charge. New Japanese phones offer Earthquake early warning alerts. Retrieved on 2008- 01-08. In the event of an emergency, disaster response crews can locate trapped or injured people using the signals from their mobile phones. An interactive menu accessible through the phone's Internet browser notifies the company if the user is safe or in distress. In Finland rescue services suggest hikers carry mobile phones in case of emergency even when deep in the forests beyond cellular coverage, as the radio signal of a cellphone attempting to connect to a base station can be detected by overflying rescue aircraft with special detection gear. Also, users in the United States can sign up through their provider for free text messages when an AMBER Alert goes out for a missing person in their area.

However, most mobile telephone networks operate close to capacity during normal times and spikes in call volumes caused by widespread emergencies often overload the system just when it is needed the most. Examples reported in the media where this have occurred include the 2001 September 11 attacks, the Hawaiian earthquake, the 2003 Northeast blackouts, the 2005 London Tube bombings, Hurricane Katrina, and the 2007 Minnesota bridge collapse. Thus mobile phones are better for isolated emergencies such as vehicle accidents.

Use by drivers

Mobile-phone use while driving is common but controversial. While few jurisdictions have banned motorists from using mobile phones while driving outright, some have banned or restricted drivers from using *hand-held* mobile phones while exempting phones operated in a *hands-free* fashion. Using a hand-held mobile phone while driving is an impediment to vehicle operation that can increase the risk of road traffic accidents. However, some studies have found similarly elevated accident rates among drivers using hands-free phones, suggesting that the distraction of a *telephone* conversation itself is a significant safety problem. A study done by the Transport Research Laboratory found that mobile phone users were four times more likely to be in a collision, regardless of whether the call was hands-free or not. This problem does not apply to conversations with a *passenger,* as passengers can regulate the flow of conversation according to the perceived level of danger, and also provides a second pair of eyes to spot hazards.

One phone in each hand

Applications

Mobile news services are expanding with many organizations providing "on-demand" news services by SMS. Some also provide "instant" news pushed out by SMS. Mobile telephony also facilitates activism and public journalism being explored by Reuters and Yahoo and small independent news companies such as Jasmine News in Sri Lanka. Companies like Monster are starting to offer mobile services such as job search and career advice. Consumer applications are on the rise and include everything from information guides on local activities and events to mobile coupons and discount offers one can use to save money on purchases. Even tools for creating websites for mobile phones are increasingly becoming available, e.g. Mobilemo.

The total value of mobile data services exceeds the value of paid services on the internet, and was worth 31 billion dollars in 2006 (source Informa). The largest categories of mobile services are music, picture downloads, videogaming, adult entertainment, gambling, video/TV.

Power

Mobile phones generally obtain power from batteries which can be recharged from mains power, a USB port or a cigarette lighter socket in a car. Formerly, the most common form of cell phone batteries were nickel metal-hydride, as they have a low size and weight. Lithium-Ion batteries are sometimes used, as they are lighter and do not have the voltage depression that nickel metal-hydride batteries do. Many mobile phone manufacturers have now switched to using lithium-Polymer batteries as opposed to the older Lithium-Ion, the main advantages of this being even lower weight and the possibility to make the battery a shape other than strict cuboid. Cell phone manufacturers have been experimenting with alternate power sources, including solar cells.

Features

There are significant questions as to who first invented the camera phone, as numerous other people received patents filed in the early 1990s for the device, including David M. Britz of AT&T Research in March of 1994 and Phillipe Kahn, who claims to have first invented it in 1997. The camera phone now holds 85% of the mobile phone market. Mobile phones often have features beyond sending text messages and making voice calls, including Internet browsing, music (MP3) playback, memo recording, personal organizer functions, e-mail, instant messaging, built-in cameras and camcorders, ringtones, games, radio, Push-to-Talk (PTT), infrared and Bluetooth connectivity, call registers, ability to watch streaming video or download video for later viewing, video calling and serve as a wireless modem for a PC, and soon will also serve as a console of sorts to online games and other high quality games (e.g. Final Fantasy Agito).

Tariff models

When cellular telecoms services were launched, phones and calls were very expensive and early mobile operators (carriers) decided to charge for all air time consumed by the mobile phone user. This resulted in the concept of charging callers for outbound calls and also for receiving calls. As mobile phone call charges diminished and phone adoption rates skyrocketed, more modern operators decided not to charge for incoming calls. Thus some markets have "Receiving Party Pays" models (also know as "Mobile Party Pays"), in which both outbound and received calls are charged, and other markets have "Calling Party Pays" models, by which only making calls produces costs, and receiving calls is free. An exception to this are international roaming tariffs, by which receiving calls are normally also charged.

The European market adopted a "Calling Party Pays" model throughout the GSM environment and soon various other GSM markets also started to emulate this model. As Receiving Party Pays systems have the undesired effect of phone owners keeping their phones turned off to avoid receiving unwanted calls, the total voice usage rates (and profits) in Calling Party Pays countries outperform those in Receiving Party Pays countries. Consequently, most countries previously with Receiving Party Pays models have either abandoned them or employed alternative marketing methods, such as massive voice call buckets, to avoid the problem of phone users keeping phones turned off.

In most countries today, including Syria, European Union nations, United Arab Emirates, Kazakhstan, Turkey, New Zealand, Korea, Japan, Pakistan, Argentina, Australia, Brazil, Chile, Colombia, India, Maldives, Malaysia, Peru, South Africa, Israel, Lebanon, Egypt and Jordan the person receiving a mobile phone call pays nothing. However, in Hong Kong, Canada, and the United States, one can be charged per minute. In the United States, a few carriers are beginning to offer unlimited received phone calls. For the Chinese mainland, it was reported that both of its two operators will adopt the caller-pays approach as early as January 2007.

Developing countries

In some developing countries with little telephone infrastructure, the mobile telephone is the telephony giving poor people access to medical and legal services. Cell phone use in developing countries has quadrupled in the last decade. The rise of cell phone technology in developing countries is often cited as an example of the leapfrog effect. In many remote regions in the third world went literally from having no telecommunications infrastructure to having satellite based communications systems.

Forensics and evidence

Law enforcement globally rely heavily upon mobile telephone evidence, to the extent that in the EU the "communications of every mobile telephone user are recorded". The concerns over terrorism and terrorist use of technology prompted an inquiry by the British House of Commons Home Affairs Select Committee into the use of evidence from mobile telephone devices, prompting leading mobile telephone forensic specialists to identify forensic techniques available in this area. NIST have published guidelines and procedures for the preservation, acquisition, examination, analysis, and reporting of digital information present on cell phones can be found under the NIST Publication SP800-101.

An example of criminal investigations using mobile phones is the initial location and ultimate identification of the terrorists of the 2004 Madrid train bombings. In the attacks, mobile phones had been used to detonate the bombs. However, one of the bombs failed to detonate, and the SIM card in the corresponding mobile phone gave the first serious lead about the terrorists to investigators. By tracking the whereabouts of the SIM card and correlating other mobile phones that had been registered in those areas, police were able to locate the terrorists.

Human health impacts

Since the introduction of mobile phones, concerns have been raised about the potential health impacts from regular use. As mobile phone penetrations grew past fixed landline penetration levels in 1998 in Finland and from 1999 in Sweden, Denmark and Norway, the Scandinavian health authorities have run continuous long term studies of effects of mobile phone radiation effects to humans, and in particular children. Numerous studies have reported and most studies consistently report no significant relationship between mobile phone use and health. Studies from the Institute of Cancer Research, National Cancer Institute and researchers at the Danish Institute of Cancer Epidemiology in Copenhagen for example showed no link between mobile phone use and cancer. The Danish study only covered analog mobile phone usage up through 1995, and subjects who started mobile phone usage after 1995 were counted as non-users in the study. The health concerns have grown as mobile phone penetration rates throughout Europe reached 80%–90% levels earlier in this decade and prolonged exposure studies have been carried out in almost all European countries again most reporting no effect, and the most alarming studies only reporting a possible effect. However, a study by the International Agency for Research on Cancer of 4,500 users found a borderline statistically significant link between tumor frequency on the same side of the head as the cell phone was used on and cell phone usage.

One study that reviewed the link between cellphones and sperm quality found that heavy mobile phone users (>4 hours per day) had significantly less viable sperm (WHO morphology score was less than half of the lower time cell phone users). A prospective study of 13 normal men found that significantly increasing their cell phone use (>6 hours each day for 5 days) caused a marked short-term reduction of sperm quality.

This is considered to be a thermal effect, since the testes are vulnerable to heating by RF energy because of poor circulation and heat is known to have adverse effects on male fertility. The eyes are the other part of the body known to be poor at dissipating heat. Experiments have shown that short duration exposure to very high levels of RF radiation can cause cataracts in rabbits. The non-thermal effects of RF radiation are an area of active study.

Environmental impacts

Impact of active cell phone usage

Like all high structures, cellular antenna masts pose a hazard to low flying aircraft. Towers over a certain height or towers that are close to airports or heliports are normally required to have warning lights. There have been reports that warning lights on cellular masts, TV-towers and other high structures can attract and confuse birds. US authorities estimate that millions of birds are killed near communication towers in the country each year.

An example of the way mobile phones and mobile networks have sometimes been perceived as a threat is the widely reported and later discredited claim that mobile phone masts are associated with the Colony Collapse Disorder (CCD) which has reduced bee hive numbers by up to 75% in many areas, especially near cities in the US. The Independent newspaper cited a scientific study claiming it provided evidence for the theory that mobile phone masts are a major cause in the collapse of bee populations, with controlled experiments demonstrating a rapid and catastrophic effect on individual hives near masts. Mobile phones were in fact not covered in the study, and the original researchers have since emphatically disavowed any connection between their research, mobile phones, and CCD, specifically indicating that the Independent article had misinterpreted their results and created "a horror story". While the initial claim of damage to bees was widely reported, the corrections to the story were almost non-existent in the media.

Impact of disposed phones

There are more than 500 million used mobile phones in the US sitting on shelves or in landfills, and another 125 million will be added to the shelves or landfills this year alone. The problem is growing at a rate of more than two million phones per week, putting tons of toxic waste into landfills daily. Several sites including United States top ranked TradeMyCell.com, ReCellular, and MyGreenElectronics offer to remedy to this situation by buying back and recycling cell phones from users.

Technology

Mobile phones and the network they operate under vary significantly from provider to provider, and country to country. However, all of them communicate through electromagnetic radio waves with a cell site base station, the antennas of which are usually mounted on a tower, pole or building.


Mobile phone tower

The phones have a low-power transceiver that transmits voice and data to the nearest cell sites, usually not more than 8 to 13 km (approximately 5 to 8 miles) away. When the mobile phone or data device is turned on, it registers with the mobile telephone exchange, or switch, with its unique identifiers, and will then be alerted by the mobile switch when there is an incoming telephone call. The handset constantly listens for the strongest signal being received from the surrounding base stations. As the user moves around the network, the mobile device will " handoff" to various cell sites during calls, or while waiting (idle) between calls it will reselect cell sites.

Cell sites have relatively low-power (often only one or two watts) radio transmitters which broadcast their presence and relay communications between the mobile handsets and the switch. The switch in turn connects the call to another subscriber of the same wireless service provider or to the public telephone network, which includes the networks of other wireless carriers. Many of these sites are camouflaged to blend with existing environments, particularly in scenic areas.


Cell Phone tower located in Lynnwood, WA.

The dialogue between the handset and the cell site is a stream of digital data that includes digitized audio (except for the first generation analog networks). The technology that achieves this depends on the system which the mobile phone operator has adopted. The technologies are grouped by generation. The first-generation systems started in 1979 with Japan, are all analog and include AMPS and NMT. Second-generation systems, started in 1991 in Finland, are all digital and include GSM, CDMA and TDMA. Third-generation networks, which are still being deployed, started with Japan in 2001, are all digital, and offer high-speed data access in addition to voice services and include W-CDMA (known also as UMTS), and CDMA2000 EV-DO. China will launch a third 3G technlogy on the TD-SCDMA standard. Each network operator has a unique radio frequency band.

Books about mobile communication

Since 2002, many books have been written on the social impact of mobile phones:

- Agar, Jon, *Constant Touch: A Global History of the Mobile Phone*, 2004 ISBN 1840465417
- Ahonen, Tomi, *m-Profits: Making Money with 3G Services*, 2002, ISBN 0-470-84775-1
- Ahonen, Kasper and Melkko, *3G Marketing* 2004, ISBN 0-470-85100-7
- Glotz, Peter & Bertsch, Stefan, eds. *Thumb Culture: The Meaning of Mobile Phones for Society*, 2005
- Katz, James E. & Aakhus, Mark, eds. *Perpetual Contact: Mobile Communication, Private Talk, Public Performance*, 2002
- Kavoori, Anandam & Arceneaux, Noah, eds. *The Cell Phone Reader: Essays in Social Transformation*, 2006
- Levinson, Paul, *Cellphone: The Story of the World's Most Mobile Medium, and How It Has Transformed Everything!*, 2004 ISBN 1-4039-6041-0
- Ling, Rich, *The Mobile Connection: the Cell Phone's Impact on Society*, 2004 ISBN 1558609369
- Ling, Rich and Pedersen, Per, eds. *Mobile Communications: Re-negotiation of the Social Sphere*, 2005 ISBN 1852339314
- Home page of Rich Ling
- Nyíri, Kristóf, ed. *Mobile Communication: Essays on Cognition and Community*, 2003
- Nyíri, Kristóf, ed. *Mobile Learning: Essays on Philosophy, Psychology and Education*, 2003
- Nyíri, Kristóf, ed. *Mobile Democracy: Essays on Society, Self and Politics*, 2003
- Nyíri, Kristóf, ed. *A Sense of Place: The Global and the Local in Mobile Communication*, 2005
- Nyíri, Kristóf, ed. *Mobile Understanding: The Epistemology of Ubiquitous Communication*, 2006
- Plant, Dr. Sadie, *on the mobile – the effects of mobile telephones on social and individual life*, 2001
- Rheingold, Howard, *Smart Mobs: The Next Social Revolution*, 2002 ISBN 0738208612

Terminology

Related non-mobile-phone systems

Cordless telephone (portable phone)
: Cordless phones are standard telephones with radio handsets. Unlike mobile phones, cordless phones use private base stations that are not shared between subscribers. The base station is connected to a land-line. Increasingly, with wireless local loop technologies, namely DECT, the distinction is blurred.

Professional Mobile Radio
: Advanced professional mobile radio systems can be very similar to mobile phone systems. Notably, the IDEN standard has been used as both a private trunked radio system as well as the technology for several large public providers. Similar attempts have even been made to use TETRA, the European digital PMR standard, to implement public mobile networks.

Radio phone
: This is a term which covers radios which could connect into the telephone network. These phones may not be mobile; for example, they may require a mains power supply. Also, they may require the assistance of a human operator to set up a PSTN phone call.

Satellite phone
: This type of phone communicates directly with an artificial satellite which in turn relays calls to a base station or another satellite phone. A single satellite can provide coverage to a much greater area than terrestrial base stations. Satellite phones are often used in remote areas where no mobile phone coverage exists and at sea.

Terms in various countries

Motorcycle

2008/9 Schools Wikipedia Selection. Related subjects: Road transport

A **motorcycle** or **motorbike** is a single-track, two-wheeled motor vehicle powered by an engine. Styles of motorcycles vary depending on the task for which they are designed, such as long distance travel, navigating congested urban traffic, cruising, sport and racing, or off-road conditions. In many parts of the world, motorcycles are among the least expensive and most widespread forms of motorised transport.

History

Replica of the Daimler-Maybach *Reitwagen*

An historic 1941 Crocker

The inspiration for arguably the first motorcycle was designed and built by the German inventors Gottlieb Daimler and Wilhelm Maybach in Bad Cannstatt (since 1905 a city district of Stuttgart) in 1885. The first petroleum-powered vehicle, it was essentially a motorised bicycle, although the inventors called their invention the Reitwagen ("riding car").

However, if one counts two wheels with steam propulsion as being a motorcycle, then the first one may have been American. One such machine was demonstrated at fairs and circuses in the eastern U.S. in 1867, built by Sylvester Howard Roper of Roxbury, Massachusetts.

In 1894, Hildebrand & Wolfmüller became the first motorcycle available for purchase. In the early period of motorcycle history, many producers of bicycles adapted their designs to accommodate the new internal combustion engine. As the engines became more powerful, and designs outgrew the bicycle origins, the number of motorcycle producers increased.

Until the First World War, the largest motorcycle manufacturer in the world was Indian, producing over 20,000 bikes per year. By 1920, this honour went to Harley-Davidson, with their motorcycles being sold by dealers in 67 countries, until 1928 when DKW took over as the largest manufacturer.

After the Second World War, the BSA Group became the largest producer of motorcycles in the world, producing up to 75,000 bikes a year in the 1950s. The German company NSU Motorenwerke AG held the position of largest manufacturer from 1955 until the 1970s.

A 1913 Fabrique National in-line four with shaft drive from Belgium

A pre-war Polish Sokół 1000

From the 1960s through the 1990s, small two-stroke motorcycles were popular worldwide, partly as a result of East German Walter Kaaden's engine work in the

1950s.

Today, the Japanese manufacturers, Honda, Kawasaki, Suzuki, and Yamaha dominate the motorcycle industry, although Harley-Davidson still maintains a high degree of popularity in the United States. Recent years have also seen a resurgence in the popularity of several other brands sold in the U.S. market, including BMW, KTM, Triumph, Aprilia, Moto Guzzi and Ducati.

Outside of the USA, these brands have enjoyed continued and sustained success, although Triumph, for example, has been re-incarnated from its former self into a modern world-class manufacturer. In overall numbers, however, the Chinese currently manufacture and sell more motorcycles than any other country and exports are rising. The quality of these machines is asserted to be somewhat lower than their Japanese, European and American counterparts.

Additionally, the small-capacity scooter is very popular through most of the world. The Piaggio group of Italy, for example, is one of the world's largest producers of two-wheeled vehicles. The scooter culture has, as yet, not been adopted widely in North America.

Technical aspects

Construction

Motorcycle construction is the engineering, manufacturing, and assembly of components and systems for a motorcycle which results in performance, cost and aesthetics desired by the designer. With some exceptions, construction of modern mass-produced motorcycles has standardised on a steel or aluminium frame, telescopic forks holding the front wheel, and disc brakes. A one- to eight-cylinder gasoline powered engine coupled to a manual, five- or six-speed sequential transmission drives the swingarm-mounted rear wheel by a chain, driveshaft or belt.

Fuel economy

Motorcycle fuel economy benefits from the relatively small mass of the vehicle, compared to its passengers and to other motor vehicles, and subsequent small engine displacement. However, poor aerodynamics of exposed passengers and engines designed for goals other than fuel economy can work to reduce these benefits. Riding style has a large effect on fuel economy: some riders report being able to double fuel economy by using low accelerations and lower speeds than usual, although this is the extreme case.

Fuel economy varies greatly with engine displacement and riding style ranging from a low of 29 mpg (U.S.) (8.1 l/100 km) reported by a Honda VTR1000F rider, to 107 mpg (U.S.) (2.2 l/100 km) reported for the Verucci Nitro 50 cc Scooter. A specially designed Matzu Matsuzawa Honda XL125 achieved 470 mpg (U.S.) (0.5 l/100 km) "on real highways - in real conditions."

Dynamics

Motorcycles must be leaned in order to turn. This lean can be induced by a method known as countersteering, in which the rider turns the handlebars in the direction opposite of the desired direction of turn. In other words, press forward on the handgrip in the desired direction—press right to go right, press left to go left. This concept is counter-intuitive and often very confusing to novices—and even to many experienced motorcyclists.

Short wheelbase motorcycles, such as sport bikes, can generate enough torque at the rear wheel and enough stopping force at the front wheel to lift the other wheel off the pavement. These actions, if performed on purpose, are known as wheelies and stoppies respectively. If carried past the point of recovery the resulting upset is known as "looping" the vehicle.

Additions

Various additions may be attached to a motorcycle or come as an integral part of a motorcycle from the factory.

Fairing / screen

Image:Goldwing.rally.bristol.7

Full fairing on a Honda Gold Wing

A plastic or fibreglass shell, known as a **" fairing"**, is placed over the frame on some models to shield the rider from the wind, aid in aerodynamics and protect engine components in an accident. Drag is the major factor that limits motorcycle speed, as it increases at the square of the velocity, with the resultant required power increasing with the *cube* of velocity. As can be seen from the streamlined appearance of new performance motorcycles, there is much aerodynamic technology included in the design. Unfortunately, since the 1958 FIM ban on "dustbin" fairings no major manufacturer has been bold enough to overcome the effect of the turbulence caused by the spinning front wheel, which prevents the motorcycle from cutting a clean path through the air. The "dustbin" fairing can improve aerodynamic performance without unacceptably compromising the rider's ability to control the machine, although with a weight penalty.

Screens, also called **windshields** or **windscreens**, can be built into a fairing or be attached to an otherwise unfaired bike. They are usually made from transparent high-impact acrylic plastic. They may be shaped specifically to direct air flow over or around the head of the rider even if they are much shorter than the seated rider.

An example of a fairing on a Honda CBR1000F

In the absence of a fairing or screen, a phenomenon known as the *windsock effect* occurs at speeds above 100 km/h (62 mph), where the rider becomes a major source of drag and is pushed back from the handlebars, tiring the rider. However, these motorcycles still effectively push their way through the atmosphere with brute force. A cabin cycle, which has a hull that wraps around the basic cycle frame, solved the problem of aerodynamics by isolating driver from outside air.

Modern fairings on touring and sport-touring motorcycles dramatically improve a rider's comfort and attention on long rides by reducing the effect of the wind and rain on the body. They also help keep a rider warm in cold weather or high wind chill conditions, reducing hypothermia. Heated hand grips, and even heated seats, also improve rider comfort in cold weather. Motorcycles from a number of manufacturers now have electric screens, introduced on the 1986 BMW K100LT, which raise and lower the screen with the push of a button to the optimum height for conditions.

Saddlebags or panniers


Craven's *Golden Arrow* panniers

Saddlebags or panniers mount on either side of the rear wheel behind the saddle to carry parts, tools, and/or travel gear. They can be made of fibreglass, ABS, leather, Cordura, or other appropriate sturdy material. They are normally standard items on touring motorcycles, but are usually optional on other types of motorcycles. They can be model-specific and available from a motorcycle's manufacturer, or after-market and designed to fit on numerous models.


Heated handgrips on a BMW

Heated hand grips/seats

As motorcycles lack climate control or proper protection from the wind, some manufacturers offer heated seats or hand grips to relieve the discomfort of low temperatures experienced during night riding or the colder months. They can also be added on as after market accessories and are powered by the bike's electrical system.

Luggage rack

A common addition to many bikes is an attachment onto which bags or other luggage can be fastened. This removes the need for rider backpacks and is generally a more secure and safe way to add carrying capacity to a motorcycle.

Sidecar


IMZ-Ural motorcycle with sidecar

A **sidecar** is a one-wheeled device attached to the side of a motorcycle, producing a three-wheeled vehicle. Early sidecars were removable devices that could be detached from the motorcycle. Sidecars gradually superseded forecars and trailers. The forecar comprised a two-wheeled attachment at the front of the motorcycle. The trailer was just that, pulling the passenger along behind. In neither case could rider and passenger converse easily, and early sidecars were often called 'sociable' attachments.

Trailer hitch

A **trailer hitch** or **tow hitch** is a device mounted on a motorcycle that enables it to tow a motorcycle trailer, usually to haul additional gear. No motorcycle manufacturer recommends trailer towing because it creates safety hazards for motorcyclists.

Trunk

A **motorcycle trunk** is a storage compartment in the vicinity of the seat, other than panniers or saddlebags. A trunk mounted above and rear of the seat is called a top box.

Social aspects

Subcultures


A motorcycle rally in Ontario

Around the world, motorcycles have historically been associated with subcultures. Some of these subcultures have been loose-knit social groups such as the cafe racers of 1950s Britain, and the Mods and Rockers of the 1960s. A few are believed to be criminal gangs.

Social motorcyclist organisations are popular and are sometimes organised geographically, focus on individual makes, or even specific models. Example motorcycle clubs include: American Motorcyclist Association, Harley Owners Group, Moto Guzzi National Owners Club, Gold Wing Road Riders (GWRRA), and BMW MOA.

Many motorcycle organisations raise money for charities through organised events and rides. Some organisations hold large international motorcycle rallies in different parts of the world that are attended by many thousands of riders.

Some other motorcycle organisations exist only for the direct benefit of others. Bikers Against Child Abuse (BACA) is one example. BACA assigns members to individual children to help them through difficult situations, or even stay with the child if the child is alone or frightened.

In recent decades, motorcyclists have formed political lobbying organisations in order to influence legislators to introduce motorcycle-friendly legislation. One of the oldest such organisations, the British Motorcycle Action Group, was founded in 1973 specifically in response to helmet compulsion, introduced without public consultation. In addition, the British Motorcyclists Federation (BMF), originally founded in 1960 as a reaction to the public perception of motorcyclists as leather-jacketed hooligans, has itself moved into political lobbying.

Likewise, the U.S. has ABATE, which, like most such organisations, also works to improve motorcycle safety, as well as running the usual charity fund-raising events and rallies, often for motorcycle-related political interests. Some other lobbying organisations are listed in Category:Motorcyclists organizations.

Mobility

While the reasons for people choosing to ride motorcycles are many and varied, those reasons are increasingly practical, with riders opting for a powered two-wheeler as a cost-efficient alternative to infrequent and expensive public transport systems, or as a means of avoiding or reducing the effects of urban congestion. In places where it is permitted, lane splitting, also known as filtering, allows motorcycles to use the space between vehicles to move through stationary or slow traffic.

In the UK, motorcycles are exempt from the £8 per day London congestion charge other vehicles must pay to enter the city during the day. Motorcycles are also exempt from toll charges at some river crossings, such as the Severn Bridge, Dartford Crossing, and Mersey Tunnels. Some cities, such as Bristol, allow motorcycles to use bus lanes and provide dedicated free parking. In the United States, those states that have high-occupancy vehicle lanes also allow for motorcycle travel in them. Other countries have similar policies.

In many cultures motorcycles are the primary means of motorised transport. According to the Taiwanese government, for example, "the number of automobiles per ten thousand population is around 2,500, and the number of motorcycles is about 5,000."

Safety

Motorcycles have a higher rate of fatal accidents than automobiles. United States Department of Transportation data for 2005 from the Fatality Analysis Reporting System show that for passenger cars, 18.62 fatal crashes occur per 100,000 registered vehicles. For motorcycles this figure is higher at 75.19 per 100,000 registered vehicles – four times higher than for cars. The same data show that 1.56 fatalities occur per 100 million vehicle miles travelled for passenger cars, whereas for motorcycles the figure is 43.47 – 28 times higher than for cars. Furthermore for motorcycles the accident rates have increased significantly since the end of the 1990s, while the rates have dropped for passenger cars.

The two major causes of motorcycle accidents in the United States are: motorists pulling out or turning in front of motorcyclists and violating their rights-of-way and motorcyclists running wide through turns. The former is sometimes called a SMIDSY, an acronym formed from the motorists' common response of "Sorry mate, I didn't see you". The latter is more common when motorcyclists mix drinking with riding. Motorcyclists can anticipate avoid these crashes with proper training, increasing their conspicuousness to other traffic, and separating alcohol and riding.

The United Kingdom has a number of organisations which are dedicated to improving motorcycle safety by providing advanced rider training over and above what is necessary to pass the basic motorcycle test. These include the Institute of Advanced Motorists (IAM) and the Royal Society for the Prevention of Accidents (RoSPA). Along with increased personal safety, riders with these advanced qualifications often benefit from reduced insurance costs.


An MSF rider course for novices

Motorcycle Safety Education is offered throughout the United States by a number of organisations ranging from state agencies to non-profit organisations to corporations. The courses, designed by the Motorcycle Safety Foundation (MSF), include a Basic Rider Course, an Intermediate Rider Course and an Advanced Rider Course.

In the UK and some Australian jurisdictions, such as New South Wales, the Australian Capital Territory and the Northern Territory, it is compulsory to undertake a rider training course before being issued a Learners Licence.

In Canada, motorcycle rider training is compulsory in Quebec and Manitoba only, but all provinces and territories have Graduated Licensing programs which place restrictions on new drivers until they have gained experience. Eligibility for a full motorcycle license or endorsement for completing a Motorcycle Safety course varies from province to province. The Canada Safety Council (CSC), a non-profit safety organisation, offers the Gearing Up program across Canada and is endorsed by the Motorcycle and Moped Industry Council. Training course graduates may qualify for reduced insurance premiums.

Types of motorcycles

There are three major types of motorcycle: street, off-road, and dual purpose. Within these types, there are many different sub-types of motorcycles for many different purposes.

Street

- **Choppers:** Highly customised motorcycles based on a cruiser-style frame with long rake (longer front forks) and wild paint jobs. Many are created more for show than rideability.

- **Cruisers:** A range of small to large motorcycles designed for comfort and looks with a relaxed upright or reclined seating position. They often use lots of chrome and may be highly customised.
- **Electric motorcycles:** Nearly silent, zero-emission electric motor-driven vehicles. Operating range and top speed suffer because of limitations of battery technology. Fuel cells and petroleum-electric hybrids are also under development to extend the range and improve performance of the electric motors.
- **Mini bikes:** Very small bikes designed to be simple run-around fun for both children and adults. Generally they have no hand-operated clutch or gearbox to simplify operation. Also known as Mini Motos. Not street-legal in most countries and jurisdictions. May be used for racing by all age levels.
- **Mopeds:** Small, light, inexpensive, efficient rides for getting around town. Usually started by pedalling (motorcycle + pedals = moped). Mopeds typically have an engine mouted to the frame with a chain supplying the drive force to the wheel.
- **Naked bikes/Standard/Street bikes:** Naked bikes have a riding position midway between the forward position of a sports bike and the reclined position of a cruiser. Unlike touring bikes, naked bikes often have little or no fairing (hence the title). Luggage capabilities are often an optional extra. Naked bikes are popular for commuting and other city riding as the upright riding position gives greater visibility in heavy traffic (both for the rider and to other road users) and are more comfortable than the hunched over sport bikes. Note that naked bike and standard are not fully interchangeable terms. Naked refers to the lack of bodywork and standard refers to the upright riding position.
- **Scooters:** Motorbikes with a step-through frame, generally smaller wheels than those of a traditional motorcycle and an engine mounted near the rear wheel on the swingarm. Can be ridden without straddling any part of the bike and usually features a floorboard. Available in sport, commuter, and touring models and wide variety of engine sizes from the standard 50 cc to 850 cc.

Harley-Davidson Softail Heritage Classic. A typical "cruiser" design

An Italian 125 cc Cagiva Planet.
A "standard" or "naked" motorbike

- **Sport bikes:** Fast, light, sleek motorcycles designed for maximum performance, for racing or spirited road riding. They are distinguishable by their full fairings and the rider's tipped-forward seating position. They are also called "race replicas" because of their connection to the racing category for production motorcycles known as Superbike racing, and earlier similar race series (the term arose in the 1980s). The power to weight ratio of the 900 cc+ models typically matches or exceeds one bhp of power for every one kg of mass.
  - **Racing bikes:** Motorcycles designed for circuit or road racing, including mass-production motorcycles modified for motorcycle racing or sport riding.
  - **Street customs:** Highly customised motorcycles with wild paint jobs also built for show, but constructed from a sport bike frame instead of a cruiser-style frame.


A Kawasaki ZX-7RR sport bike

- **Touring motorcycles:** Touring bikes are designed for rider and passenger comfort, luggage carrying capacity, and reliability. Cruisers, sport bikes and some dual-sports can also be used as touring bikes with the addition of aftermarket luggage and sometimes seats. Common throughout the touring market are usually large-displacement fairings and windshields (for weather and wind protection), large-capacity fuel tanks (for long-range travel), engines optimised for progressive torque rather than highest possible power, and a more relaxed, basically upright seating position.
  - **Sport touring motorcycles:** Sport-tourers combine attributes of a sport bike and a touring motorcycle. They are built for comfortable long-distance travel while maintaining a forward-leaning riding position, good handling, and high performance.
- **Underbones:** Small motorcycle which is a crossover between a scooter and a true motorcycle with step-through frame, popular in Southeast Asia. While the fuel tank for most motorcycles are tear-shaped and located at the top and just behind the instrument panel, the fuel tank for an underbone motorcycle is located under the seat.


A BMW R1200RT touring bike

Off-road

- **Motocross bikes:** Motorcycles designed for racing over closed circuits, often with jumps, over varied terrain of gravel/mud/sand. Sometimes simply called "dirt bikes" when not being raced, they can also be used for informal off-road recreation, or "mudding".
- **Supermotos:** Beginning in the mid-1990s, motocross machines fitted with street wheels and tyres similar to those used on Sport bikes began to appear. These are known as "Supermotards", and riders of these machines compete in specially organised rallies and races.
- **Trials motorcycles:** Motorcycles made as light as possible, with no seat (as they are designed to be ridden standing up), in order to provide maximum freedom of body positioning and stunt capability for use in observed trials competition.


A Honda motocross bike

Dual-purpose

- **Dual-sports:** Road-legal machines offering a compromise in highway and off-road performance, durability and comfort. Since the requirements are often conflicting, the manufacturer has to choose one or the other, resulting in a great variety of bikes in this category.
- **Enduros:** Road-legal versions of a motocross machine, i.e., featuring high ground clearance and copious suspension with minimal creature comforts. Highly unsuitable for long distance road travel. The features that differ from the motocross versions are the silencers, the flywheel weights and the presence of features necessary for highway use.
- **Adventure Touring:** Closely related to dual-sports, adventure tourers are motorcycles with lighter weight than just about any other bike considered a tourer, but heavier than any traditional dual-sport. Adventure tourers can handle with aplomb rough dirt paths such as fire roads however, for their weight they are generally not suited for anything more strenuous than that. The advantage is their increased number of luxury features and larger engines which make on-road riding much more enjoyable.

The popular Kawasaki KLR650 dual-purpose motorcycle

Motorcycle rider postures

The motorcyclist's riding position depends on the geometry of the rider's body combined with the geometry of the motorcycle itself, falling along a spectrum of three basic postures:

- **Standard**: In this position the rider sits roughly upright, in a neutral position, neither leaning forward nor rearward, knees lower than the hips, and feet roughly below the riders centre of gravity. The rider has excellent visibility and a higher seat height—but with greater wind resistance, a higher centre of gravity, and potentially more difficulty **flat-footing**—having the ability, when stationary, to put both feet flat on the ground for safety and comfort, keeping the machine upright.

- **Sport**: In this position the rider leans forward with the upper torso, supporting the upper-body weight with the back, stomach and leg muscles thereby keeping the forearms loose and relaxed providing smooth steering input/feedback at the handlebars. Knees are at hip height or below and squeezed against the tank to help support the upper body with the feet positioned on the balls of the foot on the footpegs. The position offers the advantage of decreased wind resistance but an otherwise cramped position that may be difficult to sustain for longer periods (some hours). The Sport riding position offers good flat-footing.

- **Cruiser**: In this position the rider sits at a lower seat-height with the upper torso upright to slightly rearward. Knees are near hip height and legs extended forward. This position offers the advantage of comfortable circulation to the legs and ease of flat-footing—though with a lower field of visibility. In this position the rider may have difficulty lifting off the seat (when crossing an obstacle for example).

Important factors of a motorcycle's ergonomic geometry that determine the seating posture include the height, angle and location of footpegs, seat and handlebars. Likewise, factors in a rider's physical geometry that contribute to seating posture include torso, arm, thigh and leg length, and overall rider height.

Legal definitions and restrictions

A motorcycle is broadly defined by law in some countries for the purposes of registration, taxation or licensing riders as a two-wheel motor vehicle "fit to drive." Other countries distinguish between mopeds and other small bikes and the larger, more powerful vehicles. In Canada and some U.S. jurisdictions, three-wheeled motor vehicles fall under the auspices of motorcycle regulations.


In some jurisdictions, the term "motorcycle" includes trikes

In the United Kingdom, the rules on which motorcycle may be ridden by whom are complex. A "*moped*", which can be ridden at age 16, has a maximum design speed not exceeding 50 km/h (31 mph) and engine capacity no greater than 50 cc. A "*learner motorcycle*", which can be ridden from age 17, has an engine up to 125 cc with a power output not exceeding 11 kW (14.8 hp). Only a Compulsory Basic Training (CBT) license is needed to ride a learner motorcycle. A "*large restricted motorcycle*", which has a power output of not more than 25 kW (33.5 hp). Riders are restricted to riding large restricted motorcycles or smaller for two years after passing their initial motorcycle test. A "*large motorcycle*", which has a power output of at least 25 kW (33.5 hp).

For riders over age 21 there is a direct access route to gaining a licence to ride a large motorcycle, which allows somebody with no motorcycle experience to train and pass a test in around five days. All motorcycle riders in the UK must first take a one-day CBT course, regardless of which class of motorcycle they intend to ride. In addition a theory test must be taken prior to taking a practical test for any type of motorcycle licence.

In New Zealand, "learner" and "restricted" motorcycles may only have a 250 cc engine capacity. This distinction draws some criticism, as it allows 15-year-old learner riders to operate bikes capable of reaching speeds in excess of 250 km/h (155 mph).


A scooter and a motorcycle

The laws of some countries allow anyone with a car licence to legally ride mopeds not exceeding 50 cc in capacity, meaning that they do not need to show any competency in handling such a vehicle.

The laws and regulations for legal moped usage in the U.S. vary by state. The specifics of the motorcycle and moped laws in the U.S. can be obtained from each individual state's Department of Motor Vehicles' websites.

Nuclear power

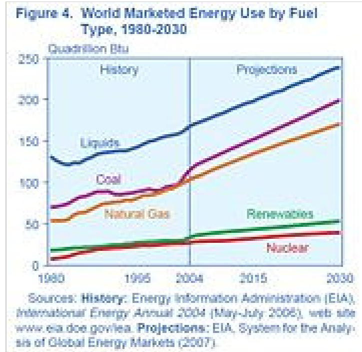
2008/9 Schools Wikipedia Selection. Related subjects: Engineering; Environment

Nuclear power is a type of nuclear technology involving the controlled use of nuclear reactions, usually nuclear fission, to release energy for work including propulsion, heat, and the generation of electricity. Nuclear energy is produced by a controlled nuclear chain reaction and creates heat—which is used to boil water, produce steam, and drive a steam turbine. The turbine can be used for mechanical work and also to generate electricity.

The Ikata Nuclear Power Plant, a pressurized water reactor that has no cooling tower, but cools by direct exchange with the ocean.

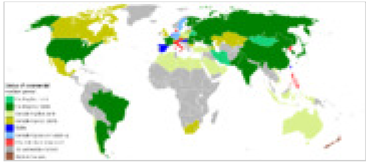
The Susquehanna Steam Electric Station, a boiling water reactor. The nuclear reactors are located inside the rectangular containment buildings towards the front of the cooling towers. The towers in the background vent water vapor.

Use



Historical and projected world energy use by energy source, 1980-2030, Source: International Energy Outlook 2007, EIA.

As of 2004, nuclear power provided 6.5% of the world's energy and 15.7% of the world's electricity, with the U.S., France, and Japan together accounting for 57% of nuclear generated electricity. As of 2007, the IAEA reported there are 439 nuclear power reactors in operation in the world, operating in 31 countries.



The status of nuclear power globally. Nations in dark green have reactors and are constructing new reactors, those in light green are constructing their first reactor, those in dark yellow are considering new reactors, those in light yellow are considering their first reactor, those in blue have reactors but are not constructing or decommissioning, those in light blue are considering decommissioning and those in red have decommissioned all their commercial reactors. Brown indicates that the country has declared itself free of nuclear power and weapons.

The United States produces the most nuclear energy, with nuclear power providing 20% of the electricity it consumes, while France produces the highest percentage of its electrical energy from nuclear reactors—80% as of 2006. In the European Union as a whole, nuclear energy provides 30% of the electricity. Nuclear energy policy differs between European Union countries, and some, such as Austria and Ireland, have no active nuclear power stations. In comparison, France has a large number of these plants, with 16 multi-unit stations in current use.

Many military and some civilian (such as some icebreaker) ships use nuclear marine propulsion, a form of nuclear propulsion.

International research is continuing into safety improvements such as passively safe plants, the use of nuclear fusion, and additional uses of process heat such as the hydrogen production (in support of a hydrogen economy), for desalinating sea water, and for use in district heating systems.

History

Origins

Nuclear fission was first experimentally achieved by Enrico Fermi in 1934 when his team bombarded uranium with neutrons. In 1938, German chemists Otto Hahn and Fritz Strassmann, along with Austrian physicists Lise Meitner and Meitner's nephew, Otto Robert Frisch, conducted experiments with the products of neutron-bombarded uranium. They determined that the relatively tiny neutron split the nucleus of the massive uranium atoms into two roughly equal pieces, which was a surprising result. Numerous scientists, including Leo Szilard who was one of the first, recognized that if fission reactions released additional neutrons, a self-sustaining nuclear chain reaction could result. This spurred scientists in many countries (including the United States, the United Kingdom, France, Germany, and the Soviet Union) to petition their government for support of nuclear fission research.

In the United States, where Fermi and Szilard had both emigrated, this led to the creation of the first man-made reactor, known as Chicago Pile-1, which achieved criticality on December 2, 1942. This work became part of the Manhattan Project, which built large reactors at the Hanford Site (formerly the town of Hanford, Washington) to breed plutonium for use in the first nuclear weapons. A parallel uranium enrichment effort also was pursued.

After World War II, the fear that reactor research would encourage the rapid spread of nuclear weapons and technology, combined with what many scientists thought would be a long road of development, created a situation in which reactor research was kept under strict government control and classification. In addition, most reactor research centered on purely military purposes.

Electricity was generated for the first time by a nuclear reactor on December 20, 1951 at the EBR-I experimental station near Arco, Idaho, which initially produced about 100 kW (the Arco Reactor was also the first to experience partial meltdown, in 1955). In 1952, a report by the Paley Commission (*The President's Materials Policy Commission*) for President Harry Truman made a "relatively pessimistic" assessment of nuclear power, and called for "aggressive research in the whole field of solar energy." A December 1953 speech by President Dwight Eisenhower, " Atoms for Peace," emphasized the useful harnessing of the atom and set the U.S. on a course of strong government support for international use of nuclear power.

Early years

In 1954, Lewis Strauss, then chairman of the United States Atomic Energy Commission (forerunner of the U.S. Nuclear Regulatory Commission and the United States Department of Energy) spoke of electricity in the future being "too cheap to meter." While few doubt he was thinking of atomic energy when he made the statement, he may have been referring to hydrogen fusion, rather than uranium fission. Actually, the consensus of government and business at the time was that nuclear (fission) power might eventually become merely economically competitive with conventional power sources.

The Shippingport Atomic Power Station in Shippingport, Pennsylvania was the first commercial reactor in the USA and was opened in 1957.

On June 27, 1954, the USSRs Obninsk Nuclear Power Plant became the world's first nuclear power plant to generate electricity for a power grid, and produced around 5 megawatts electric power.

In 1955 the United Nations' "First Geneva Conference", then the world's largest gathering of scientists and engineers, met to explore the technology. In 1957 EURATOM was launched alongside the European Economic Community (the latter is now the European Union). The same year also saw the launch of the International Atomic Energy Agency (IAEA).

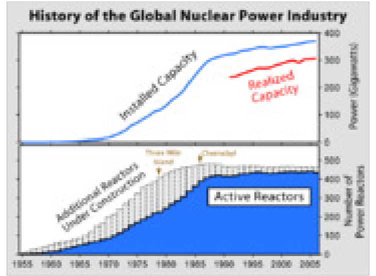
The world's first commercial nuclear power station, Calder Hall in Sellafield, England was opened in 1956 with an initial capacity of 50 MW (later 200 MW). The first commercial nuclear generator to become operational in the United States was the Shippingport Reactor (Pennsylvania, December, 1957).

One of the first organizations to develop nuclear power was the U.S. Navy, for the purpose of propelling submarines and aircraft carriers. It has a good record in nuclear safety, perhaps because of the stringent demands of Admiral Hyman G. Rickover, who was the driving force behind nuclear marine propulsion as well as the Shippingport Reactor. The U.S. Navy has operated more nuclear reactors than any other entity, including the Soviet Navy, with no publicly known major incidents. The first nuclear-powered submarine, USS *Nautilus* (SSN-571), was put to sea in December 1954. Two U.S. nuclear submarines, USS *Scorpion* and *Thresher*, have been lost at sea. These vessels were both lost due to malfunctions in systems not related to the reactor plants. Also, the sites are monitored and no known leakage has occurred from the onboard reactors.

Enrico Fermi and Leó Szilárd in 1955 shared for the nuclear reactor, belatedly granted for the work they had done during the Manhattan Project.

Development

Installed nuclear capacity initially rose relatively quickly, rising from less than 1 gigawatt (GW) in 1960 to 100 GW in the late 1970s, and 300 GW in the late 1980s. Since the late 1980s worldwide capacity has risen much more slowly, reaching 366 GW in 2005. Between around 1970 and 1990, more than 50 GW of capacity was under construction (peaking at over 150 GW in the late 70s and early 80s) — in 2005, around 25 GW of new capacity was planned. More than two-thirds of all nuclear plants ordered after January 1970 were eventually cancelled.



History of the use of nuclear power (top) and the number of active nuclear power plants (bottom).

During the 1970s and 1980s rising economic costs (related to extended construction times largely due to regulatory changes and pressure-group litigation) and falling fossil fuel prices made nuclear power plants then under construction less attractive. In the 1980s (U.S.) and 1990s (Europe), flat load growth and electricity liberalization also made the addition of large new baseload capacity unattractive.

The 1973 oil crisis had a significant effect on countries, such as France and Japan, which had relied more heavily on oil for electric generation (39% and 73% respectively) to invest in nuclear power. Today, nuclear power supplies about 80% and 30% of the electricity in those countries, respectively.


Washington Public Power Supply System Nuclear Power Plants 3 and 5 were never completed.

A general movement against nuclear power arose during the last third of the 20th century, based on the fear of a possible nuclear accident, fears of radiation, nuclear proliferation, and on the opposition to nuclear waste production, transport and final storage. Perceived risks on the citizens' health and safety, the 1979 accident at Three Mile Island and the 1986 Chernobyl disaster played a part in stopping new plant construction in many countries, although the public policy organization Brookings Institution suggests that new nuclear units have not been ordered in the U.S. because the Institution's research concludes they cost 15–30% more over their lifetime than conventional coal and natural gas fired plants.

Unlike the Three Mile Island accident, the much more serious Chernobyl accident did not increase regulations affecting Western reactors since the Chernobyl reactors were of the problematic RBMK design only used in the Soviet Union, for example lacking "robust" containment buildings. Many of these reactors are still in use today. However, changes were made in both the reactors themselves (use of low enriched uranium) and in the control system (prevention of disabling safety systems) to prevent the possibility of a duplicate accident.

An international organization to promote safety awareness and professional development on operators in nuclear facilities was created: WANO; World Association of Nuclear Operators.

Opposition in Ireland, New Zealand and Poland prevented nuclear programs there, while Austria (1978), Sweden (1980) and Italy (1987) (influenced by Chernobyl) voted in referendums to oppose or phase out nuclear power.

Future of the industry

As of 2007, Watts Bar 1, which came on-line in Feb. 7, 1996, was the last U.S. commercial nuclear reactor to go on-line. This is often quoted as evidence of a successful worldwide campaign for nuclear power phase-out. However, political resistance to nuclear power has only ever been successful in New Zealand, and parts of Europe and the Philippines. Even in the U.S. and throughout Europe, investment in research and in the nuclear fuel cycle has continued, and some experts predict that electricity shortages, fossil fuel price increases, global warming and heavy metal emissions from fossil fuel use, new technology such as passively safe plants, and national energy security will renew the demand for nuclear power plants.

Many countries remain active in developing nuclear power, including Japan, China and India, all actively developing both fast and thermal technology, South Korea and the United States, developing thermal technology only, and South Africa and China, developing versions of the Pebble Bed Modular Reactor (PBMR). Several EU member states actively pursue nuclear programs, while some other member states continue to have a ban for the nuclear energy use. Finland has a new European Pressurized Reactor under construction by Areva, which is currently two years behind schedule. On December 20, 2002 the Bulgarian Council of Ministers voted to restart construction of the Belene Nuclear Power Plant. The plant's foundations were laid in 1987, however construction was abandoned in 1990, with the first reactor being 40% ready. It is expected that the first reactor should go on-line in 2013, and the second in 2014.

Japan has an active nuclear construction program with new units brought on-line in 2005. In the U.S., three consortia responded in 2004 to the U.S. Department of Energy's solicitation under the Nuclear Power 2010 Program and were awarded matching funds—the Energy Policy Act of 2005 authorized loan guarantees for up to six new reactors, and authorized the Department of Energy to build a reactor based on the Generation IV Very-High-Temperature Reactor concept to produce both electricity and hydrogen. As of the early 21st century, nuclear power is of particular interest to both China and India to serve their rapidly growing economies—both are developing fast breeder reactors. See also energy development. In the energy policy of the United Kingdom it is recognized that there is a likely future energy supply shortfall, which may have to be filled by either new nuclear plant construction or maintaining existing plants beyond their programmed lifetime.

On September 22, 2005 it was announced that two sites in the U.S. had been selected to receive new power reactors (exclusive of the new power reactor scheduled for INL). In August 2007, TVA was approved to restart construction of Watts Bar 2. The reactor is scheduled to be completed and come online in 2013. Currently, no new reactors have been ordered in the United States. However, as of February 2008, five applications for Combined Licenses (COL) have been submitted . Note that these applications are not declarations of intent to build new power plants, but submission of a COL application is one of the final steps a utility must take before construction can begin on a new nuclear reactors.

Russia has begun building floating nuclear power plants. The £100 million ($204.9 million, 2 billion руб) vessel, the *Lomonosov*, to be completed in 2010, is the first of seven plants that Moscow says will bring vital energy resources to remote Russian regions. While producing only a small fraction of the power of a standard Russian land-based plant, it can supply power to a city of 200,000, or function as a desalination plant. The Russian atomic energy agency said that at least 12 countries were also interested in buying floating nuclear plants.

In January 2008, the United Kingdom confirmed a new generation of nuclear power plants to be built in order to meet the country's growing energy crisis. The government hopes that the first station will be operational before 2020.

There is a possible impediment to production of nuclear power plants, due to a backlog at Japan Steel Works, the only factory in the world able to manufacture the central part of a nuclear reactor's containment vessel in a single piece, which reduces the risk of a radiation leak. The company can only make four per year of the steel forgings, which contain radioactivity in a nuclear reactor. It will double its capacity in the next two years, but still will not be able to to meet current global ddemand promptly. Utilities across the world are submitting orders years in advance of any actual need. Other manufacturers are examining various options, including making the component themselves, or finding ways to make a similar item using alternate methods.

Nuclear reactor technology

Cattenom Nuclear Power Plant.

Conventional thermal power plants all have a fuel source to provide heat. Examples are gas, coal, or oil. For a nuclear power plant, this heat is provided by nuclear fission inside the nuclear reactor. When a relatively large fissile atomic nucleus is struck by a neutron it forms two or more smaller nuclei as fission products, releasing energy and neutrons in a process called nuclear fission. The neutrons then trigger further fission, and so on. When this nuclear chain reaction is controlled, the energy released can be used to heat water, produce steam and drive a turbine that generates electricity. While a nuclear power plant uses the same fuel, uranium-235 or plutonium-239, a nuclear explosive involves an uncontrolled chain reaction, and the rate of fission in a reactor is not capable of reaching sufficient levels to trigger a nuclear explosion because commercial reactor grade nuclear fuel is not enriched to a high enough level. Naturally found uranium contains 0.711% U-235 by mass, the rest being U-238 and trace amounts of other isotopes. Most reactor fuel is enriched to only 3–4%, but some designs use natural uranium or highly enriched uranium. Reactors for nuclear submarines and large naval surface ships, such as aircraft carriers, commonly use highly enriched uranium. Although highly enriched uranium is more expensive, it reduces the frequency of refueling, which is very useful for military vessels. CANDU reactors are able to use unenriched uranium because the heavy water they use as a moderator and coolant does not absorb neutrons like light water does.

The chain reaction is controlled through the use of materials that absorb and moderate neutrons. In uranium-fueled reactors, neutrons must be moderated (slowed down) because slow neutrons are more likely to cause fission when colliding with a uranium-235 nucleus. Light water reactors use ordinary water to moderate and cool the reactors. When at operating temperatures if the temperature of the water increases, its density drops, and fewer neutrons passing through it are slowed enough to trigger further reactions. That negative feedback stabilizes the reaction rate.

The current types of plants (and their common components) are discussed in the article nuclear reactor technology.

A number of other designs for nuclear power generation, the Generation IV reactors, are the subject of active research and may be used for practical power generation in the future. A number of the advanced nuclear reactor designs could also make critical fission reactors much cleaner, much safer and/or much less of a risk to the proliferation of nuclear weapons.

It should be noted that such Geneation IV reactors are not necessarily fuel by uranium but by thorium, a more abundent fertile material that decays into U233 after being exposed to neutrons. Such reactors use about 1/300 the amount of fuel to power them. The Liquid Fluoride Reactor is one such example of this.

Life cycle

A nuclear reactor is only part of the life-cycle for nuclear power. The process starts with mining. Generally, uranium mines are either open-pit strip mines, or in-situ leach mines. In either case, the uranium ore is extracted, usually converted into a stable and compact form such as yellowcake, and then transported to a processing facility. Here, the yellowcake is converted to uranium hexafluoride, which is then enriched using various techniques. At this point, the enriched uranium, containing more than the natural 0.7% U-235, is used to make rods of the proper composition and geometry for the particular reactor that the fuel is destined for. The fuel rods will spend about 3 operational cycles (typically 6 years total now) inside the reactor, generally until about 3% of their uranium has been fissioned, then they will be moved to a spent fuel pool where the short lived isotopes generated by fission can decay away. After about 5 years in a cooling pond, the spent fuel is radioactively and thermally cool enough to handle, and it can be moved to dry storage casks or reprocessed.

The Nuclear Fuel Cycle begins when uranium is mined, enriched, and manufactured into nuclear fuel, (1) which is delivered to a nuclear power plant. After usage in the power plant, the spent fuel is delivered to a reprocessing plant (2) or to a final repository (3) for geological disposition. In reprocessing 95% of spent fuel can be recycled to be returned to usage in a power plant (4).

Water

Like all forms of power generation using steam turbines, Nuclear power plants use copious amounts of water for cooling. As with most power plants, two-thirds of the energy produced by a nuclear power plant goes into waste heat (see Carnot cycle), and that heat is carried away from the plant in the water (which remains uncontaminated by radioactivity). The emitted water either is sent into cooling towers where it goes up and is emitted as water droplets (literally a cloud) or is discharged into large bodies of water - cooling ponds, lakes, rivers, or oceans. Droughts can pose a severe problem by causing the source of cooling water to run out.

The Palo Verde Nuclear Generating Station near Phoenix, AZ is the only nuclear generating facility in the world that is not located adjacent to a large body of water. Instead, it uses treated sewage from several nearby municipalities to meet its cooling water needs, recycling 20 billion US gallons (76,000,000 m^3) of wastewater each year.

Like conventional power plants, nuclear power plants generate large quantities of waste heat which is expelled in the condenser, following the turbine. Collocation of plants that can take advantage of this thermal energy has been suggested by Oak Ridge National Laboratory (ORNL) as a way to take advantage of process synergy for added energy efficiency. One example would be to use the power plant steam to produce hydrogen from water. The hydrogen would cost less, and the nuclear power plant would exhaust less heat into the atmosphere and water vapor (which is a greenhouse gas).

Solid waste

The safe storage and disposal of nuclear waste is a significant challenge. The most important waste stream from nuclear power plants is spent fuel. A large nuclear reactor produces 3 cubic metres (25–30 tonnes) of spent fuel each year. It is primarily composed of unconverted uranium as well as significant quantities of transuranic actinides (plutonium and curium, mostly). In addition, about 3% of it is made of fission products. The actinides (uranium, plutonium, and curium) are responsible for the bulk of the long term radioactivity, whereas the fission products are responsible for the bulk of the short term radioactivity.

High level radioactive waste

Spent fuel is highly radioactive and needs to be handled with great care and forethought. However, spent nuclear fuel becomes less radioactive over time. After 40 years, the radiation flux is 99.9% lower than it was the moment the spent fuel was removed, although still dangerously radioactive.

Spent fuel rods are stored in shielded basins of water (spent fuel pools), usually located on-site. The water provides both cooling for the still-decaying fission products, and shielding from the continuing radioactivity. After a few decades some on-site storage involves moving the now cooler, less radioactive fuel to a dry-storage facility or dry cask storage, where the fuel is stored in steel and concrete containers until its radioactivity decreases naturally ("decays") to levels safe enough for other processing. This interim stage spans years or decades, depending on the type of fuel. Most U.S. waste is currently stored in temporary storage sites requiring oversight, while suitable permanent disposal methods are discussed.

As of 2003, the United States had accumulated about 49,000 metric tons of spent nuclear fuel from nuclear reactors. Underground storage at Yucca Mountain in U.S. has been proposed as permanent storage. After 10,000 years of radioactive decay, according to United States Environmental Protection Agency standards, the spent nuclear fuel will no longer pose a threat to public health and safety.

The amount of waste can be reduced in several ways, particularly reprocessing. Even so, the remaining waste will be substantially radioactive for at least 300 years even if the actinides are removed, and for up to thousands of years if the actinides are left in. Even with separation of all actinides, and using fast breeder reactors to destroy by transmutation some of the longer-lived non-actinides as well, the waste must be segregated from the environment for one to a few hundred years, and therefore this is properly categorized as a long-term problem. Subcritical reactors or fusion reactors could also reduce the time the waste has to be stored. It has been argued that the best solution for the nuclear waste is above ground temporary storage since technology is rapidly changing. The current waste may well become a valuable resource in the future.

In the U.S., which does not reprocess nuclear waste, one source said "Already more than 80,000 tonnes of highly radioactive waste sits in cooling pools next to the 103 US nuclear power plants, awaiting transportation to a storage facility yet to be found. This dangerous material will be an attractive target for terrorist sabotage as it travels through 39 states on roads and railway lines for the next 25 years". Even keeping track of it all has proved to be a problem. In fact fears have been expressed that terrorists could gain control of some of it to make " dirty bombs" or, if reprocessing were ever instituted in the U.S., perhaps even a nuclear device.

France is one of the world's most densely populated countries. According to a 2007 story broadcast on *60 Minutes*, nuclear power gives France the cleanest air of any industrialized country, and the cheapest electricity in all of Europe. France reprocesses its nuclear waste to reduce its mass and make more energy. However, the article continues, "Today we stock containers of waste because currently scientists don't know how to reduce or eliminate the toxicity, but maybe in 100 years perhaps scientists will ... Nuclear waste is an enormously difficult political problem which to date no country has solved. It is, in a sense, the Achilles heel of the nuclear industry ... If France is unable to solve this issue, says Mandil, then 'I do not see how we can continue our nuclear program.'" Further, reprocessing itself has its critics, such as the Union of Concerned Scientists.

Low-level radioactive waste

The nuclear industry also produces a volume of low-level radioactive waste in the form of contaminated items like clothing, hand tools, water purifier resins, and (upon decommissioning) the materials of which the reactor itself is built. In the United States, the Nuclear Regulatory Commission has repeatedly attempted to allow low-level materials to be handled as normal waste: landfilled, recycled into consumer items, et cetera. Most low-level waste releases very low levels of radioactivity and is only considered radioactive waste because of its history. For example, according to the standards of the NRC, the radiation released by coffee is enough to treat it as low level waste.

Comparing radioactive waste to industrial toxic waste

In countries with nuclear power, radioactive wastes comprise less than 1% of total industrial toxic wastes, which remain hazardous indefinitely unless they decompose or are treated so that they are less toxic or, ideally, completely non-toxic. Overall, nuclear power produces far less waste material than fossil-fuel based power plants. Coal-burning plants are particularly noted for producing large amounts of toxic and mildly radioactive ash due to concentrating naturally occurring metals and radioactive material from the coal. Contrary to popular belief, coal power actually results in more radioactive waste being released into the environment than nuclear power. The population effective dose equivalent from radiation from coal plants is 100 times as much as nuclear plants.

Reprocessing

Reprocessing can potentially recover up to 95% of the remaining uranium and plutonium in spent nuclear fuel, putting it into new mixed oxide fuel. This would produce a reduction in long term radioactivity within the remaining waste, since this is largely short-lived fission products, and reduces its volume by over 90%. Reprocessing of civilian fuel from power reactors is currently done on large scale in Britain, France and (formerly) Russia, will be in China and perhaps India, and is being done on an expanding scale in Japan. The full potential of reprocessing has not been achieved because it requires breeder reactors, which are not yet commercially available. France is generally cited as the most successful reprocessor, but it presently only recycles 28% (by mass) of the yearly fuel use, 7% within France and another 21% in Russia.

Unlike other countries, the US has stopped civilian reprocessing as one part of US non-proliferation policy, since reprocessed material such as plutonium can be used in nuclear weapons. Spent fuel is all currently treated as waste. In February, 2006, a new U.S. initiative, the Global Nuclear Energy Partnership was announced. It would be an international effort to reprocess fuel in a manner making nuclear proliferation unfeasible, while making nuclear power available to developing countries.

Depleted uranium

Uranium enrichment produces many tons of depleted uranium (DU) which consists of U-238 with most of the easily fissile U-235 isotope removed. U-238 is a tough metal with several commercial uses — for example, aircraft production, radiation shielding, and making bullets and armor — as it has a higher density than lead. There are concerns that U-238 may lead to health problems in groups exposed to this material excessively, like tank crews and civilians living in areas where large quantities of DU ammunition have been used.

Debate on nuclear power

Proponents of nuclear energy aver that nuclear power is a sustainable energy source that reduces carbon emissions and increases energy security by decreasing dependence on foreign oil. Proponents also claim that the risks of storing waste are small and can be further reduced by the technology in the new reactors and the operational safety record is already good when compared to the other major kinds of power plants.

Critics claim that nuclear power is an uneconomic and potentially dangerous energy source with a limited fuel supply, and dispute whether the costs and risks can be reduced through new technology. Critics also point to the problem of storing radioactive waste, the potential for possibly severe radioactive contamination by accident or sabotage, the possibility of nuclear proliferation and the disadvantages of centralized electrical production.

Arguments of economics and safety are used by both sides of the debate.

Other issues crucial to viability and public confidence. These include long-term waste management, leaks, past meltdown near-misses and scandals, such as the Sellafield Mox site scandal reported in the Guardian as involving "the falsification of documents, which led to the resignation of John Taylor, chief executive of BNFL"

Reliability

Nuclear power plants in the U.S. now routinely reach 90% capacity factors (including planned outages), making them suitable for base load power plant operations. Nuclear plants typically strive to schedule their refuelling and maintenance outages in the spring (when hydropower is at a maximum) and to a lesser extent in the fall (both times when electricity demand is lower than the maximums in summer and winter).

The World Nuclear Association states that "Sun, wind, tides and waves cannot be controlled to provide directly either continuous base-load power, or peak-load power when it is needed. In practical terms they are therefore limited to some 10-20% of the capacity of an electricity grid, and cannot directly be applied as economic substitutes for coal or nuclear power, however important they may become in particular areas with favourable conditions." "The fundamental problem, especially for electricity supply, is their variable and diffuse nature. This means either that there must be reliable duplicate sources of electricity, or some means of electricity storage on a large scale. Apart from pumped-storage hydro systems, no such means exist at present and nor are any in sight." "Relatively few places have scope for pumped storage dams close to where the power is needed, and overall efficiency is low. Means of storing large amounts of electricity as such in giant batteries or by other means have not been developed." (Opponents dispute these claims as discussed in the main article.)

Economics

This is a controversial subject, since multi-billion dollar investments ride on the choice of an energy source.

Which power source (generally coal, natural gas, nuclear or wind) is most cost-effective depends on the assumptions used in a particular study—several are quoted in the main article.

Environmental effects

The primary environmental impacts of nuclear power include Uranium mining, radioactive effluent emissions, direct and indirect greenhouse gas emissions (water vapor, $CO2$, $NO2$) and waste heat. Which power source produces the least amount of greenhouse gases is controversial since also renewables produce indirect greenhouse emissions from sources such as mining and construction. Nuclear generation does not directly produce sulfur dioxide, nitrogen oxides, mercury or other pollutants associated with the combustion of fossil fuels.

Other issues include disposal of nuclear waste, with high level waste proposed to go in Deep geological repositories and nuclear decommissioning.

Safety

The topic of nuclear safety covers:

- The research and testing of the possible incidents/events at a nuclear power plant,
- What equipment and actions are designed to prevent those incidents/events from having serious consequences,
- The calculation of the probabilities of multiple systems and/or actions failing thus allowing serious consequences,
- The evaluation of the worst-possible timing and scope of those serious consequences (the worst-possible in extreme cases being a release of radiation),
- The actions taken to protect the public during a release of radiation,
- The training and rehearsals performed to ensure readiness in case an incident/event occurs.

Numerous different and usually redundantly duplicated safety features have been designed into (and in some cases backfitted to) nuclear power plants. In the United States, the Nuclear Regulatory Commission (NRC) has the ultimate responsibility for nuclear safety.

Accidents

The International Nuclear Event Scale (INES), developed by the International Atomic Energy Agency (IAEA), is used to communicate the severity of nuclear accidents on a scale of 0 to 7. The two most well-known events are the Three Mile Island accident and the Chernobyl disaster.

The Chernobyl disaster in 1986 at the Chernobyl Nuclear Power Plant in the Ukrainian Soviet Socialist Republic (now Ukraine) was the worst nuclear accident in history and is the only event to receive an INES score of 7. The power excursion and resulting steam explosion and fire spread radioactive contamination across large portions of Europe. The UN report 'CHERNOBYL : THE TRUE SCALE OF THE ACCIDENT' published 2005 concluded that the death toll includes the 50 workers who died of acute radiation syndrome, nine children who died from thyroid cancer, and an estimated 4000 excess cancer deaths in the future. This accident occurred due to both the flawed operation of the reactors and critical design flaws in the Soviet RBMK reactors, such as lack of a containment building. This disaster however has lead to some "lessons learned" for Western power plants, large improvements in safety at Soviet-designed nuclear power plants and major improvements to the remaining RBMK reactors.

The Mayak accident in Russia (INES 6) occurred September 29 1957, when the failure of the cooling system for a tank storing tens of thousands of tons of dissolved nuclear waste resulted in a non-nuclear explosion having a force estimated at about 75 tons of TNT.

The 1979 accident at Three Mile Island Unit 2 was the worst civilian nuclear accident outside the Soviet Union (INES score of 5). The reactor experienced a partial core meltdown. However, according to the NRC, the reactor vessel and containment building were not breached and little radiation was released to the environment, with no significant impact on health or the environment. Several studies have found no increase in cancer rates. However, a 1997 study by Dr. Steven Wing found higher cancer rates downwind of the reactor. Three scientific journals had refused to print the findings Steven Wing, and some of his fellow epidemiologists dismiss him as an anti-nuclear activist who let his personal views cloud his objectivity. The event resulted in fundamental changes in how plants in the West were to be maintained and operated.

However many point to the possibility of a catastrophic accident which could affect many thousands or even millions. Greenpeace has produced a report titled *An American Chernobyl: Nuclear "Near Misses" at U.S. Reactors Since 1986* which "reveals that nearly two hundred "near misses" to nuclear meltdowns have occurred in the United States". At almost 450 nuclear plants in the world that risk is greatly magnified, they say. This is not to mention numerous incidents, many supposedly unreported, that have occurred. Another report produced by Greenpeace called Nuclear Reactor Hazards: Ongoing Dangers of Operating Nuclear Technology in the 21st Century claims that risk of a major accident has increased in the past years.

Underlying much of the distrust is the fact that it has often been the case that populations are not informed of hazards from various technologies that may impact on them. For example Brookhaven National Laboratory's leaking of radioactive tritium into community groundwater for up to 12 years which angered the local community, dangerous coverups at the Rocky Flats Nuclear Weapons Plant or the pollution of Anniston, Alabama and other locations by Monsanto that went unreported for four decades, however such mistrust is often misdirected — while the industrial sites that were built to support the Manhattan Project and the Cold War's nuclear arms race in the United States display many cases of significant environmental contamination and other safety concerns, in the US such facilities are operated and regulated completely separately from commercial nuclear power plants.

For the future, design changes are being pursued to lessen the risks of fission reactors; in particular, passively safe plants (such as the ESBWR) are available to be built and inherently safe designs are being pursued. Fusion reactors, which may be viable in the future, have no risk of explosive radiation-releasing accidents, and even smaller risks than the already extremely small risks associated with nuclear fission. Whilst fusion power reactors will produce a very small amount of reasonably short lived, intermediate-level radioactive waste at decommissioning time, as a result of neutron activation of the reactor vessel, they will not produce any high-level, long-lived materials comparable to those produced in a fission reactor. Even this small radioactive waste aspect can be mitigated through the use of low-activation steel alloys for the tokamak vessel.

Contrasting radioactive accident emissions with industrial emissions

Claims exist that the problems of nuclear waste do not come anywhere close to approaching the problems of fossil fuel waste. A 2004 article from the BBC states: "The World Health Organization (WHO) says 3 million people are killed worldwide by outdoor air pollution annually from vehicles and industrial emissions, and 1.6 million indoors through using solid fuel." In the U.S. alone, fossil fuel waste kills 20,000 people each year. A coal power plant releases 100 times as much radiation as a nuclear power plant of the same wattage. It is estimated that during 1982, US coal burning released 155 times as much radioactivity into the atmosphere as the Three Mile Island incident. In addition, fossil fuel waste causes global warming, which leads to increased deaths from hurricanes, flooding, and other weather events.

The World Nuclear Association provides a comparison of deaths due to accidents among different forms of energy production. In their comparison, deaths per TW-yr of electricity produced from 1970 to 1992 are quoted as 885 for hydropower, 342 for coal, 85 for natural gas, and 8 for nuclear. Air pollution from fossil fuels is argued to cause tens of thousands of additional deaths each year in the US alone. Furthermore, a 2004 news article from the BBC stated, "The World Health Organization (WHO) says 3 million people are killed worldwide by outdoor air pollution annually from vehicles and industrial emissions, and 1.6 million indoors through using solid fuel. Most are in poor countries."

Health effect on population near nuclear plants

Most human exposure to radiation comes from natural background radiation. Most of the remaining exposure comes from medical procedures. Several large studies in the US, Canada, and Europe have found no evidence of any increase in cancer mortality among people living near nuclear facilities. For example, in 1991, the National Cancer Institute (NCI) of the National Institutes of Health announced that a large-scale study, which evaluated mortality from 16 types of cancer, found no increased incidence of cancer mortality for people living near 62 nuclear installations in the United States. The study showed no increase in the incidence of childhood leukemia mortality in the study of surrounding counties after start-up of the nuclear facilities. The NCI study, the broadest of its kind ever conducted, surveyed 900,000 cancer deaths in counties near nuclear facilities.

A couple of fishermen near the decommissioned Trojan Nuclear Power Plant. The reactor dome is visible on the left, and the large cooling tower on the right.

Some areas of Britain near industrial facilities, particularly near Sellafield, have displayed elevated childhood leukemia levels, in which children living locally are 10 times more likely to contract the cancer. One study of those near Sellafield has ruled out any contribution from nuclear sources, and the reasons for these increases, or clusters, are unclear. Apart from anything else, the levels of radiation at these sites are orders of magnitude too low to account for the excess incidences reported. One explanation is viruses or other infectious agents being introduced into a local community by the mass movement of migrant workers. Likewise, small studies have found an increased incidence of childhood leukemia near some nuclear power plants has been found in Germany and France. Nonetheless, the results of larger multi-site studies in these countries invalidate the hypothesis of an increased risk of leukemia related to nuclear discharge. The methodology and very small samples in the studies finding an increased incidence has been criticized.

In December of 2007, it was reported that a study showed that German children who lived near nuclear power plants had a higher rate of cancer than those who did not. However, the study also stated that there was no extra radiation near the nuclear power plants, and scientists were puzzled as to what was causing the higher rate of cancer.

Nuclear proliferation and terrorism concerns

Nuclear proliferation is the spread of nuclear weapons and related technology to nations not recognized as "Nuclear Weapon States" by the Nuclear Nonproliferation Treaty. Since the days of the Manhattan Project it has been known that reactors could be used for weapons-development purposes—the first nuclear reactors were developed for exactly this reason—as the operation of a nuclear reactor converts U-238 into plutonium. As a consequence, since the 1950s there have been concerns about the possibility of using reactors as a dual-use technology, whereby apparently peaceful technological development could serve as an approach to nuclear weapons capability.

Original impetus for development of nuclear power came from the military nuclear programs, including the early designs of power reactors that were developed for nuclear submarines. In many countries nuclear and civilian nuclear programs are linked, at least by common research projects and through agencies such as the U.S. DOE. In the U.S., for example, the first goal of the Department of Energy is "to advance the national, economic, and energy security of the United States; to promote scientific and technological innovation in support of that mission; and to ensure the environmental cleanup of the national nuclear weapons complex."

To prevent weapons proliferation, safeguards on nuclear technology were published in the Nuclear Non-Proliferation Treaty (NPT) and monitored since 1968 by the International Atomic Energy Agency (IAEA). Nations signing the treaty are required to report to the IAEA what nuclear materials they hold and their location. They agree to accept visits by IAEA auditors and inspectors to verify independently their material reports and physically inspect the nuclear materials concerned to confirm physical inventories of them in exchange for access to nuclear materials and equipment on the global market.

Several states did not sign the treaty and were able to use international nuclear technology (often procured for civilian purposes) to develop nuclear weapons (India, Pakistan, Israel, and South Africa). Of those who have signed the treaty and received shipments of nuclear paraphernalia, many states have either claimed to, or been accused of, attempting to use supposedly civilian nuclear power plants for developing weapons. Certain types of reactors may be more conducive to producing nuclear weapons materials than others, such as possible future fast breeder reactors, and a number of international disputes over proliferation have centered on the specific model of reactor being contracted for in a country suspected of nuclear weapon ambitions.

There is concern in some countries over North Korea and Iran operating research reactors and fuel enrichment plant. In 2006, North Korea detonated what they claimed was a functioning nuclear weapon, which analysis has indicated was fueled by plutonium, presumably diverted from their Yongbyon nuclear reactor. North Korea has since signed a deal with the United States regarding its Yongbyon plant and has discontinued its nuclear activities. An IAEA report also recently cited "significant cooperation" by Iran and that it has slowed its enrichment of uranium. See also Nuclear program of Iran.

Aside from their plutonium-producing potential, some research reactors are considered proliferation threats because of their use of highly-enriched uranium (HEU) as their fuel. According to the IAEA, there are over 100 reactors in the world which continue to be fueled by HEU, though for decades work has pursued to convert these to operate with low-enriched uranium (LEU). In this case, the threat is not considered to be based on surrepticious weapons development, but rather that of theft of the enriched nuclear materials, which would help potential bomb makers subvert the largest hurdle in developing an enriched-uranium weapon.

Vulnerability of plants to attack

Nuclear power plants are generally (although not always) considered "hard" targets. In the US, plants are surrounded by a double row of tall fences which are electronically monitored. The plant grounds are patrolled by a sizeable force of armed guards. The NRC's "Design Basis Threat" criteria for plants is a secret, and so what size attacking force the plants are able to protect against is unknown. However, to scram a plant takes less than 5 seconds while unimpeded restart takes hours, severely hampering a terrorist force in a goal to release radioactivity.

Attack from the air is a more problematic concern. The most important barrier against the release of radioactivity in the event of an aircraft strike is the containment building and its missile shield. The NRC's Chairman has said "Nuclear power plants are inherently robust structures that our studies show provide adequate protection in a hypothetical attack by an airplane. The NRC has also taken actions that require nuclear power plant operators to be able to manage large fires or explosions—no matter what has caused them."

In addition, supporters point to large studies carried out by the US Electric Power Research Institute that tested the robustness of both reactor and waste fuel storage, and found that they should be able to sustain a terrorist attack comparable to the September 11 terrorist attacks in the USA. Spent fuel is usually housed inside the plant's "protected zone" or a spent nuclear fuel shipping cask; stealing it for use in a " dirty bomb" is extremely difficult. Exposure to the intense radiation would almost certainly quickly incapacitate or kill any terrorists who attempt to do so.

Nuclear power plants are designed to withstand threats deemed credible at the time of licensing. However, as weapons evolve it cannot be said unequivocably that within the 60 year life of a plant it will not become vulnerable. In addition, the future status of storage sites may be in doubt. Other forms of energy production are also vulnerable to attack, such as hydroelectric dams and LNG tankers.

Use of waste byproduct as a weapon

An additional concern with nuclear power plants is that if the by-products of nuclear fission—the nuclear waste generated by the plant—were to be unprotected it could be used as a radiological weapon, colloquially known as a " dirty bomb". There have been incidents of nuclear plant workers attempting to sell nuclear materials for this purpose (for example, there was such an incident in Russia in 1999 where plant workers attempted to sell 5 grams of radioactive material on the open market, and an incident in 1993 where Russian workers were caught attempting to sell 4.5 kilograms of enriched uranium.), and there are additional concerns that the transportation of nuclear waste along roadways or railways opens it up for potential theft. The UN has since called upon world leaders to improve security in order to prevent radioactive material falling into the hands of terrorists, and such fears have been used as justifications for centralized, permanent, and secure waste repositories and increased security along transportation routes.

Optical fibre

2008/9 Schools Wikipedia Selection. Related subjects: Engineering; Physics


Optical fibers

An **optical fibre** (or **fibre**) is a glass or plastic fibre that carries light along its length. **Fibre optics** is the overlap of applied science and engineering concerned with the design and application of optical fibers. Optical fibers are widely used in fibre-optic communication, which permits transmission over longer distances and at higher data rates than other forms of communications. Fibers are used instead of metal wires because signals travel along them with less loss, and they are immune to electromagnetic interference. Optical fibers are also used to form sensors, and in a variety of other applications.

Light is kept in the "core" of the optical fibre by total internal reflection. This causes the fibre to act as a waveguide. Fibers which support many propagation paths or transverse modes are called multimode fibers (MMF). Fibers which support only a single mode are called singlemode fibers (SMF). Multimode fibers generally have a large-diameter core, and are used for short-distance communication links or for applications where high power must be transmitted. Singlemode fibers are used for most communication links longer than 200 meters.

Joining lengths of optical fibre is more complex than joining electrical wire or cable. The ends of the fibers must be carefully cleaved, and then spliced together either mechanically or by fusing them together with an electric arc. Special connectors are used to make removable connections.

History

Guiding of light by refraction, the principle that makes fibre optics possible, was first demonstrated by Daniel Colladon and Jacques Babinet in Paris in the 1840s, with Irish inventor John Tyndall offering public displays using water-fountains ten years later. Practical applications, such as close internal illumination during dentistry, appeared early in the twentieth century. Image transmission through tubes was demonstrated independently by the radio experimenter Clarence Hansell and the television pioneer John Logie Baird in the 1920s. The principle was first used for internal medical examinations by Heinrich Lamm in the following decade. In 1952, physicist Narinder Singh Kapany conducted experiments that led to the invention of optical fiber, based on Tyndall's earlier studies; modern optical fibers, where the glass fibre is coated with a transparent cladding to offer a more suitable refractive index, appeared later in the decade. Development then focused on fiber bundles for image transmission. The first fibre optic semi-flexible gastroscope was patented by Basil Hirschowitz, C. Wilbur Peters, and Lawrence E. Curtiss, researchers at the University of Michigan, in 1956. In the process of developing the gastroscope, Curtiss produced the first glass-clad fibers; previous optical fibers had relied on air or impractical oils and waxes as the low-index cladding material. A variety of other image transmission applications soon followed. The advent of ultrapure silicon for semiconductor devices made low-loss silica fibre practical.

In 1965, Charles K. Kao and George A. Hockham of the British company Standard Telephones and Cables were the first to suggest that attenuation of contemporary fibers was caused by impurities, which could be removed, rather than fundamental physical effects such as scattering. They speculated that optical fibre could be a practical medium for communication, if the attenuation could be reduced below 20 dB per kilometer. This attenuation level was first achieved in 1970, by researchers Robert D. Maurer, Donald Keck, Peter C. Schultz, and Frank Zimar working for American glass maker Corning Glass Works, now Corning Inc. They demonstrated a fibre with 17 dB optic attenuation per kilometer by doping silica glass with titanium. A few years later they produced a fibre with only 4 dB/km using germanium oxide as the core dopant. Such low attenuations ushered in optical fiber telecommunications and enabled the Internet. Nowadays, attenuations in optical cables are far less than those in electrical copper cables, leading to long-haul fibre connections with repeater distances of 500–800 km.

The erbium-doped fibre amplifier, which reduced the cost of long-distance fibre systems by reducing or even in many cases eliminating the need for optical-electrical-optical repeaters, was co-developed by teams led by David Payne of the University of Southampton, and Emmanuel Desurvire at Bell Laboratories in 1986. The more robust optical fibre commonly used today utilizes glass for both core and sheath and is therefore less prone to aging processes. It was invented by Gerhard Bernsee in 1973 by Schott Glass in Germany.

In 1991, the emerging field of photonic crystals led to the development of photonic crystal fibre (*Science* (2003), vol 299, page 358), which guides light by means of diffraction from a periodic structure, rather than total internal reflection. The first photonic crystal fibers became commercially available in 1996. Photonic crystal fibers can be designed to carry higher power than conventional fibre, and their wavelength dependent properties can be manipulated to improve their performance in certain applications.

Applications

Optical fibre communication

Optical fibre can be used as a medium for telecommunication and networking because it is flexible and can be bundled as cables. It is especially advantageous for long-distance communications, because light propagates through the fibre with little attenuation compared to electrical cables. This allows long distances to be spanned with few repeaters. Additionally, the light signals propagating in the fibre can be modulated at rates as high as 40 Gb/s , and each fibre can carry many independent channels, each by a different wavelength of light (wavelength-division multiplexing). Over short distances, such as networking within a building, fiber saves space in cable ducts because a single fiber can carry much more data than a single electrical cable. Fibre is also immune to electrical interference, which prevents cross-talk between signals in different cables and pickup of environmental noise. Also, wiretapping is more difficult compared to electrical connections, and there are concentric dual core fibers that are said to be tap-proof. Because they are non-electrical, fibre cables can bridge very high electrical potential differences and can be used in environments where explosive fumes are present, without danger of ignition.

Although fibers can be made out of transparent plastic, glass, or a combination of the two, the fibers used in long-distance telecommunications applications are always glass, because of the lower optical attenuation. Both multi-mode and single-mode fibers are used in communications, with multi-mode fiber used mostly for short distances (up to 500 m), and single-mode fibre used for longer distance *links*. Because of the tighter tolerances required to couple light into and between single-mode fibers (core diameter about 10 micrometers), single-mode transmitters, receivers, amplifiers and other components are generally more expensive than multi-mode components.

Fibre optic sensors

Optical fibers can be used as sensors to measure strain, temperature, pressure and other parameters. The small size and the fact that no electrical power is needed at the remote location gives the fibre optic sensor an advantage over a conventional electrical sensor in certain applications.

Optical fibers are used as hydrophones for seismic or SONAR applications. Hydrophone systems with more than 100 sensors per fibre cable have been developed. Hydrophone sensor systems are used by the oil industry as well as a few countries' navies. Both bottom mounted hydrophone arrays and towed streamer systems are in use. The German company Sennheiser developed a microphone working with a laser and optical fibers.

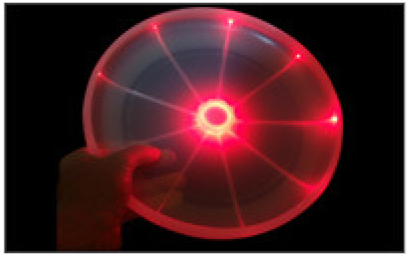
Optical fiber sensors for temperature and pressure have been developed for downhole measurement in oil wells. The fibre optic sensor is well suited for this environment as it is functioning at temperatures too high for semiconductor sensors (Distributed Temperature Sensing).

Another use of the optical fibre as a sensor is the optical gyroscope which is in use in the Boeing 767 and in some car models (for navigation purposes) and the use in Hydrogen microsensors.

Fibre-optic sensors have been developed to measure co-located temperature and strain simultaneously with very high accuracy. This is particularly useful when acquiring information from small complex structures.

Other uses of optical fibers

Fibers are widely used in illumination applications. They are used as light guides in medical and other applications where bright light needs to be shone on a target without a clear line-of-sight path. In some buildings, optical fibers are used to route sunlight from the roof to other parts of the building (see non-imaging optics). Optical fibre illumination is also used for decorative applications, including signs, art, and artificial Christmas trees. Swarovski boutiques use optical fibers to illuminate their crystal showcases from many different angles while only employing one light source. Optical fibre is an intrinsic part of the light-transmitting concrete building product, LiTraCon.



A frisbee illuminated by fibre optics


A fibre-optic Christmas Tree

Optical fibre is also used in imaging optics. A coherent bundle of fibers is used, sometimes along with lenses, for a long, thin imaging device called an endoscope, which is used to view objects through a small hole. Medical endoscopes are used for minimally invasive exploratory or surgical procedures (endoscopy). Industrial endoscopes (see fiberscope or borescope) are used for inspecting anything hard to reach, such as jet engine interiors.

An optical fibre doped with certain rare-earth elements such as erbium can be used as the gain medium of a laser or optical amplifier. Rare-earth doped optical fibers can be used to provide signal amplification by splicing a short section of doped fiber into a regular (undoped) optical fiber line. The doped fibre is optically pumped with a second laser wavelength that is coupled into the line in addition to the signal wave. Both wavelengths of light are transmitted through the doped fibre, which transfers energy from the second pump wavelength to the signal wave. The process that causes the amplification is stimulated emission.

Optical fibers doped with a wavelength shifter are used to collect scintillation light in physics experiments.

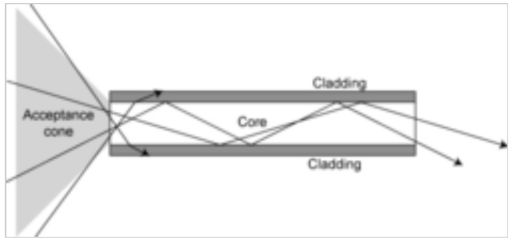
Optical fibre can be used to supply a low level of power (around one watt) to electronics situated in a difficult electrical environment. Examples of this are electronics in high-powered antenna elements and measurement devices used in high voltage transmission equipment.

Optical fibers are also used in fibre optic gyroscopes, and other interferometry instruments.

Principle of operation

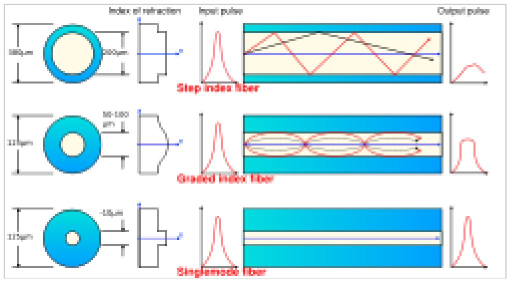
An optical fibre is a cylindrical dielectric waveguide that transmits light along its axis, by the process of total internal reflection. The fibre consists of a *core* surrounded by a cladding layer. To confine the optical signal in the core, the refractive index of the core must be greater than that of the cladding. The boundary between the core and cladding may either be abrupt, in *step-index fibre*, or gradual, in *graded-index fibre*.

Multimode fibre



The propagation of light through a multi-mode optical fibre.

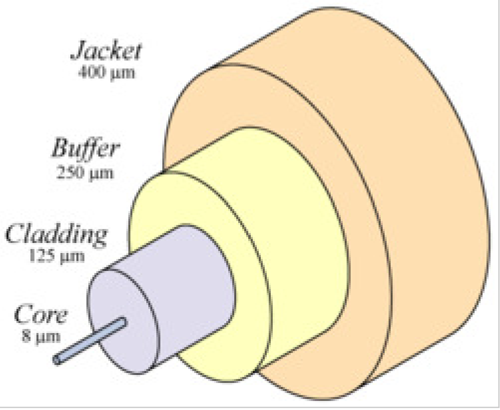
Fibre with large (greater than 10 µm) core diameter may be analyzed by geometric optics. Such fibre is called *multimode fibre*, from the electromagnetic analysis (see below). In a step-index multimode fibre, rays of light are guided along the fibre core by total internal reflection. Rays that meet the core-cladding boundary at a high angle (measured relative to a line normal to the boundary), greater than the critical angle for this boundary, are completely reflected. The critical angle (minimum angle for total internal reflection) is determined by the difference in index of refraction between the core and cladding materials. Rays that meet the boundary at a low angle are refracted from the core into the cladding, and do not convey light and hence information along the fibre. The critical angle determines the acceptance angle of the fibre, often reported as a numerical aperture. A high numerical aperture allows light to propagate down the fiber in rays both close to the axis and at various angles, allowing efficient coupling of light into the fibre. However, this high numerical aperture increases the amount of dispersion as rays at different angles have different path lengths and therefore take different times to traverse the fibre. A low numerical aperture may therefore be desirable.



Optical fibre types.

In graded-index fiber, the index of refraction in the core decreases continuously between the axis and the cladding. This causes light rays to bend smoothly as they approach the cladding, rather than reflecting abruptly from the core-cladding boundary. The resulting curved paths reduce multi-path dispersion because high angle rays pass more through the lower-index periphery of the core, rather than the high-index center. The index profile is chosen to minimize the difference in axial propagation speeds of the various rays in the fibre. This ideal index profile is very close to a parabolic relationship between the index and the distance from the axis.

Singlemode fibre



A typical single-mode optical fibre, showing diameters of the component layers.

Fibre with a core diameter less than about ten times the wavelength of the propagating light cannot be modeled using geometric optics. Instead, it must be analyzed as an electromagnetic structure, by solution of Maxwell's equations as reduced to the electromagnetic wave equation. The electromagnetic analysis may also be required to understand behaviors such as speckle that occur when coherent light propagates in multi-mode fiber. As an optical waveguide, the fibre supports one or more confined transverse modes by which light can propagate along the fiber. Fibre supporting only one mode is called single-mode or *mono-mode* fiber. The behavior of larger-core multimode fiber can also be modeled using the wave equation, which shows that such fiber supports more than one mode of propagation (hence the name). The results of such modeling of multi-mode fiber approximately agree with the predictions of geometric optics, if the fibre core is large enough to support more than a few modes.

The waveguide analysis shows that the light energy in the fibre is not completely confined in the core. Instead, especially in single-mode fibers, a significant fraction of the energy in the bound mode travels in the cladding as an evanescent wave.

The most common type of single-mode fibre has a core diameter of 8 to 10 µm and is designed for use in the near infrared. The mode structure depends on the wavelength of the light used, so that this fiber actually supports a small number of additional modes at visible wavelengths. Multi-mode fibre, by comparison, is manufactured with core diameters as small as 50 micrometres and as large as hundreds of micrometres.

Special-purpose fibre

Some special-purpose optical fibre is constructed with a non-cylindrical core and/or cladding layer, usually with an elliptical or rectangular cross-section. These include polarization-maintaining fibre and fibre designed to suppress whispering gallery mode propagation.

Photonic crystal fibre is made with a regular pattern of index variation (often in the form of cylindrical holes that run along the length of the fiber). Such fibre uses diffraction effects instead of or in addition to total internal reflection, to confine light to the fiber's core. The properties of the fibre can be tailored to a wide variety of applications.

Manufacturing

Materials

Glass optical fibers are almost always made from silica, but some other materials, such as fluorozirconate, fluoroaluminate, and chalcogenide glasses, are used for longer-wavelength infrared applications. Like other glasses, these glasses have a refractive index of about 1.5. Typically the difference between core and cladding is less than one percent.

Plastic optical fibers (POF) are commonly step-index multimode fibers with a core diameter of 0.5 mm or larger. POF typically have higher attenuation co-efficients than glass fibers, 1 dB/m or higher, and this high attenuation limits the range of POF-based systems.

Process

Standard optical fibers are made by first constructing a large-diameter *preform*, with a carefully controlled refractive index profile, and then *pulling* the preform to form the long, thin optical fibre. The preform is commonly made by three chemical vapor deposition methods: *inside vapor deposition*, *outside vapor deposition*, and *vapor axial deposition*.

With *inside vapor deposition*, a hollow glass tube approximately 40 cm in length known as a "preform" is placed horizontally and rotated slowly on a lathe, and gases such as silicon tetrachloride ($SiCl_4$) or germanium tetrachloride ($GeCl_4$) are injected with oxygen in the end of the tube. The gases are then heated by means of an external hydrogen burner, bringing the temperature of the gas up to 1900 kelvins, where the tetrachlorides react with oxygen to produce silica or germania (germanium oxide) particles. When the reaction conditions are chosen to allow this reaction to occur in the gas phase throughout the tube volume, in contrast to earlier techniques where the reaction occurred only on the glass surface, this technique is called *modified chemical vapor deposition*.

The oxide particles then agglomerate to form large particle chains, which subsequently deposit on the walls of the tube as soot. The deposition is due to the large difference in temperature between the gas core and the wall causing the gas to push the particles outwards (this is known as thermophoresis). The torch is then traversed up and down the length of the tube to deposit the material evenly. After the torch has reached the end of the tube, it is then brought back to the beginning of the tube and the deposited particles are then melted to form a solid layer. This process is repeated until a sufficient amount of material has been deposited. For each layer the composition can be modified by varying the gas composition, resulting in precise control of the finished fibre's optical properties.

In outside vapor deposition or vapor axial deposition, the glass is formed by *flame hydrolysis*, a reaction in which silicon tetrachloride and germanium tetrachloride are oxidized by reaction with water (H_2O) in an oxyhydrogen flame. In outside vapor deposition the glass is deposited onto a solid rod, which is removed before further processing. In vapor axial deposition, a short *seed rod* is used, and a porous preform, whose length is not limited by the size of the source rod, is built up on its end. The porous preform is consolidated into a transparent, solid preform by heating to about 1800 kelvins.

The preform, however constructed, is then placed in a device known as a drawing tower, where the preform tip is heated and the optic fiber is pulled out as a string. By measuring the resultant fiber width, the tension on the fiber can be controlled to maintain the fibre thickness.

Practical issues

Optical fibre cables

In practical fibers, the cladding is usually coated with a tough resin *buffer* layer, which may be further surrounded by a *jacket* layer, usually plastic. These layers add strength to the fiber but do not contribute to its optical wave guide properties. Rigid fiber assemblies sometimes put light-absorbing ("dark") glass between the fibers, to prevent light that leaks out of one fibre from entering another. This reduces cross-talk between the fibers, or reduces flare in fibre bundle imaging applications.

Modern cables come in a wide variety of sheathings and armor, designed for applications such as direct burial in trenches, dual use as power lines , installation in conduit, lashing to aerial telephone poles, submarine installation, or insertion in paved streets. In recent years the cost of small fibre-count pole-mounted cables has greatly decreased due to the high Japanese and South Korean demand for fibre to the home (FTTH) installations.

Fiber cable can be very flexible, but traditional fiber's loss increases greatly if the fibre is bent with a radius smaller than around 30 mm. This creates a problem when the cable is bent around corners or wound around a spool, making FTTX installations more complicated. "Bendable fibers", targeted towards easier installation in home environments, have been standardized as ITU-T G.657. This type of fibre can be bent with a radius as low as 7.5 mm without adverse impact. Even more bendable fibers have been developed. Bendable fiber may also be resistant to fiber hacking, in which the signal in a fiber is surreptitiously monitored by bending the fibre and detecting the leakage.

Termination and splicing

Optical fibers are connected to terminal equipment by optical fibre connectors. These connectors are usually of a standard type such as *FC*, *SC*, *ST*, *LC*, or *MTRJ*.


ST fiber connector on multimode fibre

Optical fibers may be connected to each other by connectors or by *splicing*, that is, joining two fibers together to form a continuous optical waveguide. The generally accepted splicing method is arc fusion splicing, which melts the fibre ends together with an electric arc. For quicker fastening jobs, a "mechanical splice" is used.

Fusion splicing is done with a specialized instrument that typically operates as follows: The two cable ends are fastened inside a splice enclosure that will protect the splices, and the fibre ends are stripped of their protective polymer coating (as well as the more sturdy outer jacket, if present). The ends are *cleaved* (cut) with a precision cleaver to make them perpendicular, and are placed into special holders in the splicer. The splice is usually inspected via a magnified viewing screen to check the cleaves before and after the splice. The splicer uses small motors to align the end faces together, and emits a small spark between electrodes at the gap to burn off dust and moisture. Then the splicer generates a larger spark that raises the temperature above the melting point of the glass, fusing the ends together permanently. The location and energy of the spark is carefully controlled so that the molten core and cladding don't mix, and this minimizes optical loss. A splice loss estimate is measured by the splicer, by directing light through the cladding on one side and measuring the light leaking from the cladding on the other side. A splice loss under 0.1 dB is typical. The complexity of this process makes fibre splicing much more difficult than splicing copper wire.

Mechanical fiber splices are designed to be quicker and easier to install, but there is still the need for stripping, careful cleaning and precision cleaving. The fibre ends are aligned and held together by a precision-made sleeve, often using a clear index-matching gel that enhances the transmission of light across the joint. Such joints typically have higher optical loss and are less robust than fusion splices, especially if the gel is used. All splicing techniques involve the use of an enclosure into which the splice is placed for protection afterward.

Fibers are terminated in connectors so that the fiber end is held at the end face precisely and securely. A fiber-optic connector is basically a rigid cylindrical barrel surrounded by a sleeve that holds the barrel in its mating socket. The mating mechanism can be "push and click", "turn and latch" ("bayonet"), or screw-in (threaded). A typical connector is installed by preparing the fiber end and inserting it into the rear of the connector body. Quick-set adhesive is usually used so the fibre is held securely, and a strain relief is secured to the rear. Once the adhesive has set, the fiber's end is polished to a mirror finish. Various polish profiles are used, depending on the type of fiber and the application. For singlemode fiber, the fiber ends are typically polished with a slight curvature, such that when the connectors are mated the fibers touch only at their cores. This is known as a "physical contact" (PC) polish. The curved surface may be polished at an angle, to make an "angled physical contact" (APC) connection. Such connections have higher loss than PC connections, but greatly reduced back reflection, because light that reflects from the angled surface leaks out of the fibre core; the resulting loss in signal strength is known as gap loss. APC fibre ends have low back reflection even when disconnected.

Free-space coupling

It often becomes necessary to align an optical fiber with another optical fibre or an optical device such as a light-emitting diode, a laser diode, or an optoelectronic device such as a modulator. This can involve either carefully aligning the fibre and placing it in contact with the device to which it is to couple, or can use a lens to allow coupling over an air gap. In some cases the end of the fibre is polished into a curved form that is designed to allow it to act as a lens.

In a laboratory environment, the fiber end is usually aligned to the device or other fiber with a fibre launch system that uses a microscope objective lens to focus the light down to a fine point. A precision translation stage (micro-positioning table) is used to move the lens, fibre, or device to allow the coupling efficiency to be optimized.

Fibre fuse

At high optical intensities, above 2 megawatts per square centimetre, when a fibre is subjected to a shock or is otherwise suddenly damaged, a *fibre fuse* can occur. The reflection from the damage vaporizes the fibre immediately before the break, and this new defect remains reflective so that the damage propagates back toward the transmitter at 1–3 meters per second." The open fibre control system, which ensures laser eye safety in the event of a broken fiber, can also effectively halt propagation of the fibre fuse. In situations, such as undersea cables, where high power levels might be used without the need for open fiber control, a "fibre fuse" protection device at the transmitter can break the circuit to prevent any damage.

Photovoltaic array

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

A photovoltaic array is a linked assembly of PV modules.

Timber framed house with a photovoltaic array

A **photovoltaic array** is a linked collection of photovoltaic modules, which are in turn made of multiple interconnected solar cells. The cells convert solar energy into direct current electricity via the photovoltaic effect. The power that one module can produce is seldom enough to meet requirements of a home or a business, so the modules are linked together to form an *array*. Most PV arrays use an inverter to convert the DC power produced by the modules into alternating current that can plug into the existing infrastructure to power lights, motors, and other loads. The modules in a PV array are usually first connected in series to obtain the desired voltage; the individual strings are then connected in parallel to allow the system to produce more current. Solar arrays are typically measured by the electrical power they produce, in watts, kilowatts, or even megawatts.


The solar panels on this small yacht at sea can charge the 12 volt batteries at up to 9 amperes in full, direct sunlight.

Applications

In urban and suburban areas, photovoltaic arrays are commonly used on rooftops to measure power use; often the building will have a preexisting connection to the power grid, in which case the energy produced by the PV array will be sold back to the utility in some sort of net metering agreement. In more rural areas, ground-mounted PV systems are more common. The systems may also be equipped with a battery backup system to compensate for a potentially unreliable power grid. In agricultural settings, the array may be used to directly power DC pumps, without the need for an inverter. In remote settings such as mountainous areas, islands, or other places where a power grid is unavailable, solar arrays can be used as the sole source of electricity, usually by charging a storage battery. Satellites use solar arrays for their power. In particular the International Space Station uses multiple solar arrays to power all the equipment on board. Solar photovoltaic panels are frequently applied in satellite power. However, costs of production have been reduced in recent years for more widespread use through production and technological advances. For example, single crystal silicon solar cells have largely been replaced by less expensive multicrystalline silicon solar cells, and thin film silicon solar cells have also been developed recently at lower costs of production yet (see Solar cell). Although they are reduced in energy conversion efficiency from single crystalline Si wafers, they are also much easier to produce at comparably lower costs. Together with a storage battery, photovoltaics have become commonplace for certain low-power applications, such as signal buoys or devices in remote areas or simply where connection to the electricity mains would be impractical. In experimental form they have even been used to power automobiles in races such as the World solar challenge across Australia. Many yachts and land vehicles use them to charge on-board batteries.

PV performance

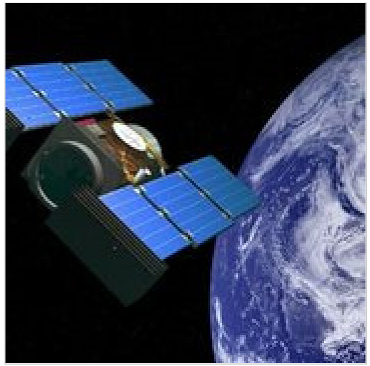
At high noon on a cloudless day at the equator, the power of the sun is about 1 kW/m², on the Earth's surface, to a plane that is perpendicular to the sun's rays. As such, PV arrays can track the sun through each day to greatly enhance energy collection. However, tracking devices add cost, and require maintenance, so it is more common for PV arrays to have fixed mounts that tilt the array and face due South in the Northern Hemisphere (in the Southern Hemisphere, they should point due North). The tilt angle, from horizontal, can be varied for season, but if fixed, should be set to give optimal array output during the peak electrical demand portion of a typical year. For large systems, the energy gained by using tracking systems outweighs the added complexity (trackers can increase efficiency by 30% or more). PV arrays that approach or exceed one megawatt often use solar trackers. Accounting for clouds, and the fact that most of the world is not on the equator, and that the sun sets in the evening, the correct measure of solar power is insolation – the average number of kilowatt-hours per square meter per day. For the weather and latitudes of the United States and Europe, typical insolation ranges from 4 kWh/m²/day in northern climes to 6.5 kWh/m²/day in the sunniest regions. Typical solar panels have an average efficiency of 12%, with the best commercially available panels at 20%. Thus, a photovoltaic installation in the southern latitudes of Europe or the United States may expect to produce 1 kWh/m²/day. A typical "150 watt" solar panel is about a square meter in size. Such a panel may be expected to produce 1 kWh every day, on average, after taking into account the weather and the latitude. In the Sahara desert, with less cloud cover and a better solar angle, one can obtain closer to 8.3 kWh/m²/day. The unpopulated area of the Sahara desert is over 9 million km², which if covered with solar panels would provide 630 terawatts total power. The Earth's current energy consumption rate is around 13.5 TW at any given moment (including oil, gas, coal, nuclear, and hydroelectric).


A solar panel on top of a parking meter. Note that this particular installation is shaded, and may not perform as desired.

Other factors affect PV performance. Photovoltaic cells' electrical output is extremely sensitive to shading. When even a small portion of a cell, module, or array is shaded, while the remainder is in sunlight, the output falls dramatically due to internal 'short-circuiting' (the electrons reversing course through the shaded portion of the p-n junction). Therefore it is extremely important that a PV installation is not shaded at all by trees, architectural features, flag poles, or other obstructions. Sunlight can be absorbed by dust, fallout, or other impurities at the surface of the module. This can cut down the amount of light that actually strikes the cells by as much as half. Maintaining a clean module surface will increase output performance over the life of the module. Module output and life are also degraded by increased temperature. Allowing ambient air to flow over, and if possible behind, PV modules reduces this problem. However, effective module lives are typically 25 years or more , so replacement costs should be considered as well.

Solar photovoltaic panels on spacecraft

Solar panels can be used on spacecraft, particularly when they are in the inner part of the solar system. They have been designed to pivot on spacecraft, so that they will always be in the direct path of solar rays. In order to optimize the amount of energy generated, solar panels on spacecraft can be equipped with a Fresnel lens, which concentrates sunlight. Because of these efforts to maximize electric production, and the fact that the Sun is mostly the only source of energy, the construction of solar cells on spacecraft could be one of the highest costs. When journeying to outer parts of the solar system (or beyond), nuclear reactors or radioisotope thermal generators are preferred, as the Sun's rays are too weak at such extreme distances to power a spacecraft. The ESA is researching the possibility of solar power satellites that would generate electricity in space and then beam it to Earth via laser or microwaves. In addition, solar power is being considered for use as a propulsion mechanism in lieu of chemical propulsion.

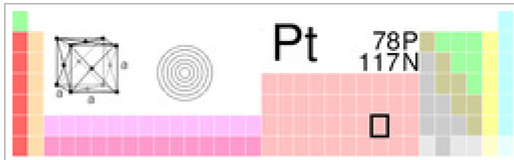


Solar panels on the Stardust spacecraft (NASA image)

Platinum

2008/9 Schools Wikipedia Selection. Related subjects: Chemical elements; Mineralogy

Platinum (pronounced /ˈplætɨnəm/) is a chemical element with the atomic symbol **Pt** and an atomic number of 78. It is in group 10 of the Periodic Table of Elements. A heavy, malleable, ductile, precious, gray-white transition metal, platinum is resistant to corrosion and occurs in some nickel and copper ores along with some native deposits. Platinum is used in jewelry, laboratory equipment, electrical contacts, dentistry, and automobile emissions control devices. Platinum bullion has the ISO currency code of XPT.

78	iridium ← platinum → gold
Pd ↑ **Pt** ↓ Ds	 Periodic Table - Extended Periodic Table

General	
Name, Symbol, Number	platinum, Pt, 78
Chemical series	transition metals
Group, Period, Block	10, 6, d
Appearance	grayish white 
Standard atomic weight	195.084 (9) $g \cdot mol^{-1}$
Electron configuration	[Xe] $4f^{14}$ $5d^{9}$ $6s^{1}$
Electrons per shell	2, 8, 18, 32, 17, 1
Physical properties	
Phase	solid
Density (near r.t.)	21.45 $g \cdot cm^{-3}$
Liquid density at m.p.	19.77 $g \cdot cm^{-3}$
Melting point	2041.4 K (1768.3 ° C, 3214.9 ° F)

Notable characteristics

When pure, the metal appears greyish-white and firm. The metal is corrosion-resistant. The catalytic properties of the six platinum family metals are outstanding. For this catalytic property, platinum is used in catalytic converters, incorporated in automobile exhaust systems, as well as tips of spark plugs.

Platinum's wear- and tarnish-resistance characteristics are well suited for making fine jewelry. Platinum is more precious than gold. The price of platinum changes along with its availability, but its price is normally more than 200% of the price of gold. In the 18th century, platinum's rarity made King Louis XV of France declare it the only metal fit for a king. Platinum possesses high resistance to chemical attack, excellent high-temperature characteristics, and stable electrical properties. All these properties have been exploited for industrial applications. Platinum does not oxidize in air at any temperature, but can be corroded by cyanides, halogens, sulfur, and caustic alkalis. This metal is insoluble in hydrochloric and nitric acid, but does dissolve in the mixture known as aqua regia (forming chloroplatinic acid). Common oxidation states of platinum include +2, and +4. The +1 and +3 oxidation states are less common, and are often stabilized by metal bonding in bimetallic (or polymetallic) species.The gold is removed from this solution as a precipitate by treatment with iron(II) chloride ($FeCl_2$). The platinum is precipitated out as impure $(NH_4)_2PtCl_6$ on treatment with NH_4Cl, leaving H_2PtCl_4 in solution.

Applications

- As a catalyst in the catalytic converter, an optional (though often mandatory by law) component of the gasoline-fueled automobile exhaust system (see "Notable characteristics" in this article).
- As a catalyst in fuel cells. Reducing the amount of platinum required (and thus cost) is a major focus of fuel cell research.
- Certain platinum-containing compounds are capable of crosslinking DNA and kill cells by similar pathways to alkylating chemotherapeutic agents. Cisplatin, carboplatin and oxaliplatin are licensed examples of this class of drugs.
- Platinum resistance thermometers.
- Electrodes for use in electrolysis and electrochemical measurements (e.g., the standard hydrogen electrode).
- In the Clark polarographic electrode for measuring oxygen tension.
- A wide range of jewelry.
- As a catalyst in the curing of silicone elastomers.
- As a catalyst in glow plugs in some model engines.
- In crucibles, alloyed with rhodium (10–40% of Rh), for high temperature melting (around 1500°C) of glass.
- In photography, it is sometimes used for archival printmaking. Platinum prints display a greater range of tones than other Black and White printing methods. Additionally platinum's chemical stability makes for extremely long-lasting prints. The disadvantage of this method, in addition to the high cost, is that platinum is less light sensitive and prints must be contact printed at the same size as the negative. Therefore, enlargements can only be made by making an enlarged negative. Platinum salts alone generally create excessive contrast in prints; combined with salts from its sister metal, palladium, produce warmer and softer tones, without diminishing the tonal range platinum enables.
- In watchmaking, Vacheron Constantin, Patek Philippe, Rolex, Breitling and other companies use platinum for producing their limited edition watch series. Watchmakers highly appreciate the unique properties of platinum as it neither tarnishes nor wears out.

Boiling point	4098 K (3825 ° C, 6917 ° F)
Heat of fusion	22.17 kJ·mol^{-1}
Heat of vaporization	469 kJ·mol^{-1}
Specific heat capacity	(25 °C) 25.86 J·mol^{-1}·K^{-1}

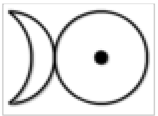
Vapor pressure

P(Pa)	1	10	100	1 k	10 k	100 k
at *T*(K)	2330	(2550)	2815	3143	3556	4094

Atomic properties	
Crystal structure	cubic face centered
Oxidation states	1, 2, 3, **4**, 5, 6 (mildly basic oxide)
Electronegativity	2.28 (Pauling scale)
Ionization energies	1st: 870 kJ/mol
	2nd: 1791 kJ/mol
Atomic radius	135 pm
Atomic radius (calc.)	177 pm
Covalent radius	128 pm
Van der Waals radius	175 pm
Miscellaneous	
Magnetic ordering	paramagnetic
Electrical resistivity	(20 °C) 105 n Ω·m
Thermal conductivity	(300 K) 71.6 W·m^{-1}·K^{-1}
Thermal expansion	(25 °C) 8.8 µm·m^{-1}·K^{-1}
Speed of sound (thin rod)	(r.t.) 2800 m·s^{-1}
Young's modulus	168 GPa

History

Naturally-occurring platinum and platinum-rich alloys have been known for a long time. Though the metal was used by pre-Columbian Native Americans, the first European reference to platinum appears in 1557 in the writings of the Italian humanist Julius Caesar Scaliger (1484– 1558) as a description of a mysterious metal found in Central American mines between Darién (Panama) and Mexico ("up until now impossible to melt by any of the Spanish arts"). The word *platinum* comes from the Spanish word *platina*, meaning "little silver."



The alchemical symbol for platinum (shown above) was made by joining the symbols of silver and gold.

Platinum was discussed by astronomer Antonio de Ulloa and Don Jorge Juan y Santacilia (1713–1773), both appointed by King Philip V to join a geographical expedition in Peru that lasted from 1735 to 1745. Amongst other things, Ulloa observed the *platina del pinto*, the unworkable metal found with gold in New Granada (Colombia). British privateers intercepted Ulloa's ship on the return voyage. Though he was well-treated in England, and even made a member of the Royal Society he was prevented from publishing a reference to the unknown metal until 1748. Before that could happen Charles Wood independently isolated the element in 1741. Major finds were discovered in Russia in 1819, which produced around 90% of the global Platinum production at the turn of the 20th century.

Due to its rarity, greater difficulty to work with and the need to alloy it with (at the time) an even more expensive metal iridium, platinum was only used in a limited way in jewelry at the end of the 19th century. This changed at beginning of the 20th century when most diamond ring mountings and most exclusive jewelry were almost completely made of platinum. From 1875 to 1960 the SI unit of length (the standard metre) was defined as the distance between two lines on a standard bar of an alloy of ninety percent platinum and ten percent iridium, measured at 0 degrees Celsius.

Occurrence

Shear modulus	61 GPa
Bulk modulus	230 GPa
Poisson ratio	0.38
Mohs hardness	4–4.5
Vickers hardness	549 MPa
Brinell hardness	392 MPa
CAS registry number	7440-06-4

Selected isotopes

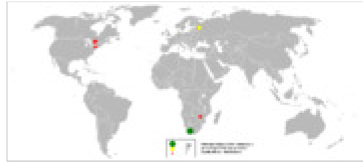
Main article: Isotopes of platinum

iso	NA	half-life	DM	DE (MeV)	DP
^{190}Pt	0.014%	6.5×10^{11} y	α	3.18	^{186}Os
^{191}Pt	syn	2.76 d	ε	?	^{191}Ir
^{192}Pt	0.782%	^{192}Pt is stable with 114 neutrons			
^{190}Pt	syn	50 y	ε	?	^{193}Ir
$^{181\,m}$Pt	syn	4.33 d	IT	0.1355 e	^{193}Pt
^{194}Pt	32.967%	^{194}Pt is stable with 116 neutrons			
^{195}Pt	33.832%	^{195}Pt is stable with 117 neutrons			
$^{195\,m}$Pt	syn	4.02 d	IT	0.1297 e	^{195}Pt
^{196}Pt	25.242%	^{196}Pt is stable with 118 neutrons			
^{197}Pt	syn	19.8913 h	β^-	0.719	^{197}Au
$^{197\,m}$Pt	syn	1.59 h	IT	0.3465	^{197}Pt
^{198}Pt	7.163%	^{198}Pt is stable with 120 neutrons			

References

Platinum ore

Platinum is an extremely rare metal, occurring as only 0.003 ppb in the Earth's crust, and is 30 times rarer than gold. If all the world's platinum reserves were poured into one Olympic-size swimming pool, it would be just deep enough to cover one's ankles. Gold would fill more than three such pools.



Platinum output in 2005

In 2005, South Africa was the top producer of platinum with almost 80% share followed by Russia and Canada, reports the British Geological Survey.

Platinum is often found chemically uncombined as native platinum and alloyed with iridium as platiniridium. The platinum arsenide, sperrylite ($PtAs_2$), is a major source of platinum associated with nickel ores in the Sudbury Basin deposit in Ontario, Canada. The rare sulfide mineral cooperite, (Pt,Pd,Ni)S, contains platinum along with palladium and nickel. Cooperite occurs in the Merensky Reef within the Bushveld complex, Gauteng, South Africa.

Platinum, often accompanied by small amounts of other platinum family metals, occurs in alluvial placer deposits in the Witwatersrand of South Africa, the Ural Mountains, and in the Absaroka Range in the American state of Montana, the only place it is found in the western hemisphere.

Platinum is produced commercially as a by-product of nickel ore processing in the Sudbury deposit. The huge quantities of nickel ore processed makes up for the fact that platinum is present as only 0.5 ppm in the ore.

Platinum exists in relatively higher abundances on the Moon and in meteorites. Corrospondingly, platinum is found in slightly higher abundances at sites of bollide impact on the Earth that are associated with resulting post-impact volcanism, and can be mined economically; the Sudbury Basin is one such example.

Precautions

According to the Centers for Disease Control and Prevention, short-term exposure to platinum salts "may cause irritation of the eyes, nose, and throat" and long-term exposure "may cause both respiratory and skin allergies." The current OSHA standard is 0.002 milligram per cubic meter of air averaged over an 8-hour work shift.

Certain platinum complexes are used in chemotherapy and show good anti-tumor activity for some tumours. Cisplatin is particularly effective against testicular cancer; cure rate was improved from 10% to 85%. However, the side effects are severe. Cisplatin causes cumulative, irreversible kidney damage and deafness.

As platinum is a catalyst in the manufacture of the silicone rubber and gel components of several types of medical implants (breast implants, joint replacement prosthetics, artificial lumbar discs, vascular access ports), the possibility that platinum free radicals could enter the body and cause adverse effects has merited study. The FDA and other countries have reviewed the issue and found no evidence to suggest toxicity in vivo.

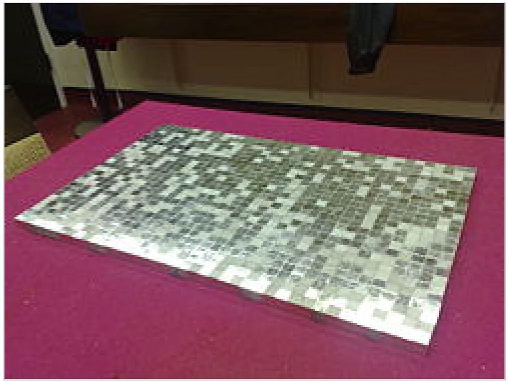
Rarity and colour

Platinum's rarity as a metal has caused advertisers to associate it with exclusivity and wealth. "Platinum" credit cards have greater privileges than do "gold" ones. "Platinum awards" are the second highest possible, ranking above gold, silver and bronze, but below "Diamond". For example, in the United States a musical album that has sold more than 1,000,000 copies, will be credited as "platinum", whereas an album that sold more than 10,000,000 copies will be certified as “diamond”. And some products, such as blenders and vehicles, with a silvery-white colour are identified as "platinum". Platinum is considered a precious metal, although its use is not as common as the use of gold or silver. The frame of the Crown of Queen Elizabeth the Queen Mother, manufactured for her Coronation as Consort of King George VI, is made of platinum. It was the first British crown to be made of this particular metal.

An assortment of native platinum nuggets

Production

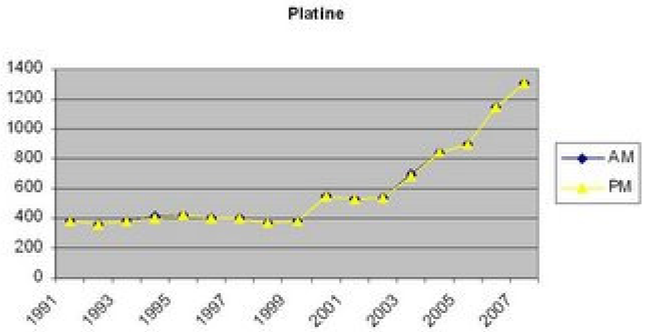
In order to obtain pure platinum, the ore is crushed, made into a slurry, and then mixed with a detergent containing "collector" molecules. Air is then blown through the mixture, enabling the grains of metal minerals to be separated from the rest of the mixture. This process is called " flotation." The next step is smelting.



1000 cubic centimetres of 99.9% pure platinum

In 2006, world supply of platinum was of about 217,700 kg or 7 million troy ounces.

Average Price from 1991 to 2007 in $ per troy ounce (~$40/g).



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Pump

2008/9 Schools Wikipedia Selection. Related subjects: Engineering


A small, electrically powered pump

A **pump** is a device used to move gases, liquids or slurries. A pump moves liquids or gases from lower pressure to higher pressure, and overcomes this difference in pressure by adding energy to the system (such as a water system). A gas pump is generally called a compressor, except in very low pressure-rise applications, such as in heating, ventilating, and air-conditioning, where the operative equipment consists of *fans* or *blowers*.

Pumps work by using mechanical forces to push the material, either by physically lifting, or by the force of compression.

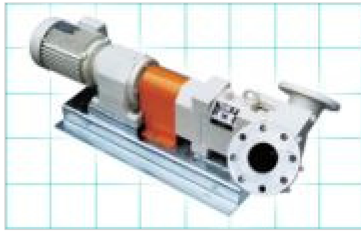
The earliest type of pump was the Archimedes screw, first used by Sennacherib, King of Assyria, for the water systems at the Hanging Gardens of Babylon and Nineveh in the 7th century BC, and later described in more detail by Archimedes in the 3rd century BC. In the 13th century AD, al-Jazari described and illustrated different types of pumps, including a reciprocating pump, double-action pump, suction pump, and piston pump.

Types


A large, electrically driven pump (electropump) for waterworks near the Hengsteysee, Germany.

Pumps fall into two major groups: **rotodynamic pumps** and **positive displacement** pumps. Their names describe the method for moving a fluid.

Positive displacement pumps


A lobe pump


Hand-operated, reciprocating, positive displacement, water pump in Košice- Ťahanovce, Slovakia (walking beam pump).

A positive displacement pump causes a fluid to move by trapping a fixed amount of it then forcing (displacing) that trapped volume into the discharge pipe. The periodic fluid displacement results in a direct increase in pressure. A positive displacement pump can be further classified as either

- a rotary-type (for example the rotary vane),
- lobe pump similar to oil pumps used in car engines, or
- the **Wendelkolben pump** or the **helical twisted Roots pump**.

Mechanism of a scroll pump

Roots-type pumps

The low pulsation rate and gentle performance of this Roots-type positive displacement pump is achieved due to a combination of its two 90° helical twisted rotors, and a triangular shaped sealing line configuration, both at the point of suction and at the point of discharge. This design produces a continuous and non-vorticuless flow with equal volume. High capacity industrial "air compressors" have been designed to employ this principle as well as most "superchargers" used on internal combustion engines.

Reciprocating-type pumps

Reciprocating-type pumps use a piston and cylinder arrangement with suction and discharge valves integrated into the pump. Pumps in this category range from having "simplex" one cylinder, to in some cases "quad" four cylinders or more. Most reciprocating-type pumps are "duplex" (two) or "triplex" (three) cylinder. Furthermore, they are either "single acting" independent suction and discharge strokes or "double acting" suction and discharge in both directions. The pumps can be powered by air, steam or through a belt drive from an engine or motor. This type of pump was used extensively in the early days of steam propulsion (19th century) as boiler feed water pumps. Though still used today, reciprocating pumps are typically used for pumping highly viscous fluids including concrete and heavy oils.

Compressed-air-powered double-diaphragm pumps

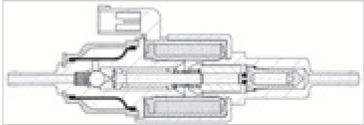
Another modern application of positive displacement pumps are compressed-air-powered double- diaphragm pumps. Run on compressed air these pumps are intrinsically safe by design, although all manufacturers offer ATEX certified models to comply with industry regulation. Commonly seen in all areas of industry from shipping to process, SandPiper, Wilden Pumps or ARO are generally the larger of the brands. They are relatively inexpensive and can be used for almost any duty from pumping water out of bunds, to pumping hydrochloric acid from secure storage (dependant on how the pump is manufactured - elastomers / body construction). Suction is normally limited to roughly 6m although heads can be almost unlimited.

Kinetic Pumps

1. Continuous energy addition
2. Conversion of added energy to increase in kinetic energy (increase in velocity)

3. Conversion of increased velocity to increase in pressure
4. Conversion of Kinetic head to Pressure Head.
5. Meet all heads like Kinetic , Potential, and Pressure

Application


metering pump for gasoline and additives

Pumps are used throughout society for a variety of purposes. Early applications includes the use of the windmill or watermill to pump water. Today, the pump is used for irrigation, water supply, gasoline supply, air conditioning systems, refrigeration (usually called a compressor), chemical movement, sewage movement, flood control, marine services, etc.

Because of the wide variety of applications, pumps have a plethora of shapes and sizes: from very large to very small, from handling gas to handling liquid, from high pressure to low pressure, and from high volume to low volume.

Liquid and slurry pumps can lose prime and this will require you to prime the pump by adding liquid to the pump and inlet pipes to get the pump started.

Specifications

Pumps are commonly rated by horsepower, flow rate, outlet pressure in feet of head, inlet suction in suction head in feet. Feet is the number of feet the pump can raise or lower a column of water at atmospheric pressure.

Pumps as public water supplies

One sort of pump once common worldwide was a hand-powered water pump over a water well where people could work it to extract water, before most houses had individual water supplies.

From this came the expression "**parish pump**" for "the sort of matter chattered about by people when they meet when they go to get water", "matter of only local interest".

Today, hand operated village pumps are considered the most sustainable low cost option for safe water supply in resource poor settings, often in rural areas in developing countries. A hand pump opens access to deeper groundwater that is often not polluted and also improves the safety of a well by protecting the water source from contaminated buckets. Pumps like the Afridev pump are designed to be cheap to build and install, and easy to maintain with simple parts. It was assumed that spare parts would become available in the local market by for-profit wholesalers. However, it became clear with time that often spare parts are not available locally, because of the low profit margins for wholesalers, especially in Africa. This means that communities are often stuck without spares and cannot use their handpump anymore and have to go back to traditional and sometimes distant, polluted resources. This is unfortunate, as water projects often have put in a lot of resources to provide that community with a handpump. As a result, spare parts free handpumps are now being developed, like the Afripump.

Radio

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Amateur Radio Station with multiple Receivers and Tranceivers

Radio is the wireless transmission of signals, by modulation of electromagnetic waves with frequencies below those of visible light.

Electromagnetic radiation travels by means of oscillating electromagnetic fields that pass through the air and the vacuum of space. It does not require a medium of transport. Information is carried by systematically changing (modulating) some property of the radiated waves, such as their amplitude or their frequency. When radio waves pass an electrical conductor, the oscillating fields induce an alternating current in the conductor. This can be detected and transformed into sound or other signals that carry information.

The word 'radio' is used to describe this phenomenon, and television and radio transmissions are classed as radio frequency emissions.

Etymology

Originally, radio or radioteleography was called 'wireless telegraphy', which was shortened to 'wireless'. The prefix *radio-* in the sense of wireless transmission was first recorded in the word *radioconductor*, coined by the French physicist Edouard Branly in 1897 and based on the verb *to radiate* (in Latin "radius" means "spoke of a wheel, beam of light, ray"). 'Radio' as a noun is said to have been coined by advertising expert Waldo Warren (White 1944). The word appears in a 1907 article by Lee de Forest, was adopted by the United States Navy in 1912 and became common by the time of the first commercial broadcasts in the United States in the 1920s. (The noun 'broadcasting' itself came from an agricultural term, meaning 'scattering seeds'.) The American term was then adopted by other languages in Europe and Asia, although British Commonwealth countries retained the term 'wireless' until the mid-20th century. In Japanese, the term 'wireless' is the basis for the term 'radio wave' although the term for the device that listens to radio waves is literally 'device for receiving sounds'.

In recent years the term 'wireless' has gained renewed popularity through the rapid growth of short range networking, e.g. WLAN ('Wireless Local Area Network'),WiFi, Bluetooth as well as mobile telephony, e.g. GSM and UMTS. Today, the term 'radio' often refers to the actual transceiver device or chip, whereas 'wireless' refers to the system and/or method used for radio communication. Hence one talks about radio transceivers and Radio Frequency Identification (RFID), but about wireless devices and wireless sensor networks.

Invention

The identity of the original inventor of radio, at the time called wireless telegraphy, is contentious. Development from a laboratory demonstration to commercial utility spanned several decades and required the efforts of many practitioners. The controversy over who invented the radio, with the benefit of hindsight, can be broken down as follows:

- In 1878, David E. Hughes transmitted Morse code by radio at and below the Super low frequency range (via a clockwork transmitter).
- In 1888, Heinrich Hertz produced and measured the Ultra High Frequency range (via a sparkgap transmitter).
- In 1891, Nikola Tesla began wireless research. He developed means to reliably produce radio frequencies, publicly demonstrated the principles of radio, and transmitted long-distance signals. He obtained a U.S. patent for the invention of the radio, as defined as "wireless transmission of data."
- Between 1893 and 1894, Roberto Landell de Moura, a Brazilian priest and scientist, conducted experiments. He did not publicise his achievement until 1900 but later obtained Brazilian and American patents.
- In 1894 in Kolkata (Calcutta), Sir Jagdish Chandra Bose (J. C. Bose) invented the Mercury Coherer (together with the telephone receiver), later used by Guglielmo Marconi to receive the radio signal in his first transatlantic radio communication over a distance of 2000 miles from Poldhu, UK to Newfoundland, St. Johns in December 1901. Guglielmo Marconi was celebrated worldwide for this achievement, but the fact that the receiver was invented by Bose was not well known.
- Alexander Stepanovich Popov, in 1894, built his first radio receiver, which contained a coherer but actually coherer was first demonstrated by J.C. Bose. Further refined as a lightning detector, he presented it to the Russian Physical and Chemical Society on May 7, 1895.
- Guglielmo Marconi was an early radio experimenter. But although frequently regarded as the true inventor of the radio, the coherer used by him was actually developed by J.C. Bose, who was ignored at the time.
- Reginald Fessenden and Lee de Forest invented amplitude-modulated (AM) radio, so that more than one station can send signals (as opposed to spark-gap radio, where one transmitter covers the entire bandwidth of the spectrum).
- Edwin H. Armstrong invented frequency-modulated (FM) radio, so that an audio signal can avoid "static," that is, interference from electrical equipment and atmospherics.

Brief history

In 1893, in St. Louis, Missouri, Tesla made devices for his experiments with electricity. Addressing the *Franklin Institute* in Philadelphia and the *National Electric Light Association*, he described and demonstrated in detail the principles of his wireless work. The descriptions contained all the elements that were later incorporated into radio systems before the development of the vacuum tube. He initially experimented with magnetic receivers, unlike the coherers (detecting devices consisting of tubes filled with iron filings which had been invented by Temistocle Calzecchi-Onesti at Fermo in Italy in 1884) used by Guglielmo Marconi and other early experimenters. .

In 1896, Marconi was awarded the British patent 12039, *Improvements in transmitting electrical impulses and signals and in apparatus there-for*, for radio. In 1897 he established the world's first radio station on the Isle of Wight, England. Marconi opened the world's first "wireless" factory in Hall Street, Chelmsford, England in 1898, employing around 50 people.

The next great invention was the vacuum tube detector, invented by the Westinghouse engineers. On Christmas Eve, 1906, Reginald Fessenden used a synchronous rotary-spark transmitter for the first radio program broadcast, from Brant Rock, Massachusetts. Ships at sea heard a broadcast that included Fessenden playing *O Holy Night* on the violin and reading a passage from the Bible. The first radio news program was broadcast August 31, 1920 by station 8MK in Detroit, Michigan. The first college radio station, 2ADD, renamed WRUC in 1940, began broadcasting October 14, 1920 from Union College, Schenectady, New York. The first regular entertainment broadcasts commenced in 1922 from the Marconi Research Centre at Writtle, near Chelmsford, England.

One of the first developments in the early 20th century (1900-1959) was that aircraft used commercial AM radio stations for navigation. This continued until the early 1960s when VOR systems finally became widespread (though AM stations are still marked on U.S. aviation charts). In the early 1930s, single sideband and frequency modulation were invented by amateur radio operators. By the end of the decade, they were established commercial modes. Radio was used to transmit pictures visible as television as early as the 1920s. Commercial television transmissions started in North America and Europe in the 1940s. In 1954, Regency introduced a pocket transistor radio, the TR-1, powered by a "standard 22.5 V Battery".

In 1960, Sony introduced its first transistorized radio, small enough to fit in a vest pocket, and able to be powered by a small battery. It was durable, because there were no tubes to burn out. Over the next 20 years, transistors replaced tubes almost completely except for very high-power uses. By 1963 colour television was being regularly transmitted commercially, and the first (radio) communication satellite, TELSTAR, was launched. In the late 1960s, the U.S. long-distance telephone network began to convert to a digital network, employing digital radios for many of its links. In the 1970s, LORAN became the premier radio navigation system. Soon, the U.S. Navy experimented with satellite navigation, culminating in the invention and launch of the GPS constellation in 1987. In the early 1990s, amateur radio experimenters began to use personal computers with audio cards to process radio signals. In 1994, the U.S. Army and DARPA launched an aggressive, successful project to construct a software radio that could become a different radio on the fly by changing software. Digital transmissions began to be applied to broadcasting in the late 1990s.

Uses of radio

Early uses were maritime, for sending telegraphic messages using Morse code between ships and land. The earliest users included the Japanese Navy scouting the Russian fleet during the Battle of Tsushima in 1905. One of the most memorable uses of marine telegraphy was during the sinking of the RMS *Titanic* in 1912, including communications between operators on the sinking ship and nearby vessels, and communications to shore stations listing the survivors.

Radio was used to pass on orders and communications between armies and navies on both sides in World War I; Germany used radio communications for diplomatic messages once its submarine cables were cut by the British. The United States passed on President Woodrow Wilson's Fourteen Points to Germany via radio during the war. Broadcasting began from San Jose in 1909 , and became feasible in the 1920s, with the widespread introduction of radio receivers, particularly in Europe and the United States. Besides broadcasting, point-to-point broadcasting, including telephone messages and relays of radio programs, became widespread in the 1920s and 1930s. Another use of radio in the pre-war years was the development of detecting and locating aircraft and ships by the use of radar (*RA*dio *D*etection *A*nd *R*anging).

Today, radio takes many forms, including wireless networks, mobile communications of all types, as well as radio broadcasting. Before the advent of television, commercial radio broadcasts included not only news and music, but dramas, comedies, variety shows, and many other forms of entertainment. Radio was unique among dramatic presentation that it used only sound. For more, see radio programming.

Audio

A Fisher 500 AM/FM hi-fi receiver from 1959.

AM broadcast radio sends music and voice in the Medium Frequency (MF—0.300 MHz to 3 MHz) radio spectrum. AM radio uses amplitude modulation, in which the amplitude of the transmitted signal is made proportional to the sound amplitude captured (transduced) by the microphone while the transmitted frequency remains unchanged. Transmissions are affected by static and interference because lightning and other sources of radio that are transmitting at the same frequency add their amplitudes to the original transmitted amplitude. The most wattage an AM radio station is allowed to use is 50,000 watts and the only stations that can blast out signals this high were grandfathered in; these include WJR and CKLW.

FM broadcast radio sends music and voice with higher fidelity than AM radio. In frequency modulation, amplitude variation at the microphone cause the transmitter frequency to fluctuate. Because the audio signal modulates the frequency and not the amplitude, an FM signal is not subject to static and interference in the same way as AM signals. FM is transmitted in the Very High Frequency (VHF—30 MHz to 300 MHz) radio spectrum. VHF radio waves act more like light, traveling in straight lines, hence the reception range is generally limited to about 50-100 miles. During unusual upper atmospheric conditions, FM signals are occasionally reflected back towards the Earth by the ionosphere, resulting in Long distance FM reception. FM receivers are subject to the capture effect, which causes the radio to only receive the strongest signal when multiple signals appear on the same frequency. FM receivers are relatively immune to lightning and spark interference.

FM Subcarrier services are secondary signals transmitted " piggyback" along with the main program. Special receivers are required to utilize these services. Analog channels may contain alternative programming, such as reading services for the blind, background music or stereo sound signals. In some extremely crowded metropolitan areas, the subchannel program might be an alternate foreign language radio program for various ethnic groups. Subcarriers can also transmit digital data, such as station identification, the current song's name, web addresses, or stock quotes. In some countries, FM radios automatically retune themselves to the same channel in a different district by using sub-bands.

Aviation voice radios use VHF AM. AM is used so that multiple stations on the same channel can be received. (Use of FM would result in stronger stations blocking out reception of weaker stations due to FM's capture effect). Aircraft fly high enough that their transmitters can be received hundreds of miles (kilometres) away, even though they are using VHF.

Marine voice radios can use AM in the shortwave High Frequency (HF—3 MHz to 30 MHz) radio spectrum for very long ranges or narrowband FM in the VHF spectrum for much shorter ranges. Government, police, fire and commercial voice services use narrowband FM on special frequencies. Fidelity is sacrificed to use a smaller range of radio frequencies, usually five kHz of deviation, rather than the 75 kHz used by FM broadcasts and 25 kHz used by TV sound.

Civil and military HF (high frequency) voice services use shortwave radio to contact ships at sea, aircraft and isolated settlements. Most use single sideband voice (SSB), which uses less bandwidth than AM. On an AM radio SSB sounds like ducks quacking. Viewed as a graph of frequency versus power, an AM signal shows power where the frequencies of the voice add and subtract with the main radio frequency. SSB cuts the bandwidth in half by suppressing the carrier and (usually) lower sideband. This also makes the transmitter about three times more powerful, because it doesn't need to transmit the unused carrier and sideband.

TETRA, Terrestrial Trunked Radio is a digital cell phone system for military, police and ambulances. Commercial services such as XM, WorldSpace and Sirius offer encrypted digital Satellite radio.

Telephony

Mobile phones transmit to a local cell site (transmitter/receiver) that ultimately connects to the public switched telephone network (PSTN) through an optic fibre or microwave radio and other network elements. When the mobile phone nears the edge of the cell site's radio coverage area, the central computer switches the phone to a new cell. Cell phones originally used FM, but now most use various digital modulation schemes. Satellite phones use satellites rather than cell towers to communicate. They come in two types: INMARSAT and Iridium. Both types provide world-wide coverage. INMARSAT uses geosynchronous satellites, with aimed high-gain antennas on the vehicles. Iridium uses 66 Low Earth Orbit satellites as the cells.

Video

Television sends the picture as AM and the sound as FM, with the sound carrier a fixed frequency (4.5 MHz in the NTSC system) away from the video carrier. Analog television also uses a vestigial sideband on the video carrier to reduce the bandwidth required.

Digital television uses quadrature amplitude modulation. A Reed-Solomon error correction code adds redundant correction codes and allows reliable reception during moderate data loss. Although many current and future codecs can be sent in the MPEG-2 transport stream container format, as of 2006 most systems use a standard-definition format almost identical to DVD: MPEG-2 video in Anamorphic widescreen and MPEG layer 2 (*MP2*) audio. High-definition television is possible simply by using a higher-resolution picture, but H.264/AVC is being considered as a replacement video codec in some regions for its improved compression. With the compression and improved modulation involved, a single "channel" can contain a high-definition program and several standard-definition programs.

Navigation

All satellite navigation systems use satellites with precision clocks. The satellite transmits its position, and the time of the transmission. The receiver listens to four satellites, and can figure its position as being on a line that is tangent to a spherical shell around each satellite, determined by the time-of-flight of the radio signals from the satellite. A computer in the receiver does the math.

Radio direction-finding is the oldest form of radio navigation. Before 1960 navigators used movable loop antennas to locate commercial AM stations near cities.

In some cases they used marine radiolocation beacons, which share a range of frequencies just above AM radio with amateur radio operators. Loran systems also used time-of-flight radio signals, but from radio stations on the ground. VOR (Very High Frequency Omnidirectional Range), systems (used by aircraft), have an antenna array that transmits two signals simultaneously. A directional signal rotates like a lighthouse at a fixed rate. When the directional signal is facing north, an omnidirectional signal pulses. By measuring the difference in phase of these two signals, an aircraft can determine its bearing or radial from the station, thus establishing a line of position. An aircraft can get readings from two VOR and locate its position at the intersection of the two radials, known as a "fix." When the VOR station is collocated with DME (Distance Measuring Equipment), the aircraft can determine its bearing and range from the station, thus providing a fix from only one ground station. Such stations are called VOR/DMEs. The military operates a similar system of navaids, called TACANs, which are often built into VOR stations. Such stations are called VORTACs. Because TACANs include distance measuring equipment, VOR/DME and VORTAC stations are identical in navigation potential to civil aircraft.

Radar

Radar (Radio Detection And Ranging) detects things at a distance by bouncing radio waves off them. The delay caused by the echo measures the distance. The direction of the beam determines the direction of the reflection. The polarization and frequency of the return can sense the type of surface. Navigational radars scan a wide area two to four times per minute. They use very short waves that reflect from earth and stone. They are common on commercial ships and long-distance commercial aircraft

General purpose radars generally use navigational radar frequencies, but modulate and polarize the pulse so the receiver can determine the type of surface of the reflector. The best general-purpose radars distinguish the rain of heavy storms, as well as land and vehicles. Some can superimpose sonar data and map data from GPS position.

Search radars scan a wide area with pulses of short radio waves. They usually scan the area two to four times a minute. Sometimes search radars use the doppler effect to separate moving vehicles from clutter. Targeting radars use the same principle as search radar but scan a much smaller area far more often, usually several times a second or more. Weather radars resemble search radars, but use radio waves with circular polarization and a wavelength to reflect from water droplets. Some weather radar use the doppler to measure wind speeds.

Emergency services

Emergency Position-Indicating Radio Beacons (EPIRBs), Emergency Locating Transmitters (ELTs) or Personal Locator Beacons (PLBs) are small radio transmitters that satellites can use to locate a person or vehicle needing rescue. Their purpose is to help rescue people in the first day, when survival is most likely. There are several types, with widely-varying performance.

Data (digital radio)

Most new radio systems are digital, see also: Digital TV, Satellite Radio, Digital Audio Broadcasting. The oldest form of digital broadcast was spark gap telegraphy, used by pioneers such as Marconi. By pressing the key, the operator could send messages in Morse code by energizing a rotating commutating spark gap. The rotating commutator produced a tone in the receiver, where a simple spark gap would produce a hiss, indistinguishable from static. Spark gap transmitters are now illegal, because their transmissions span several hundred megahertz. This is very wasteful of both radio frequencies and power.

The next advance was continuous wave telegraphy, or CW (Continuous Wave), in which a pure radio frequency, produced by a vacuum tube electronic oscillator was switched on and off by a key. A receiver with a local oscillator would " heterodyne" with the pure radio frequency, creating a whistle-like audio tone. CW uses less than 100 Hz of bandwidth. CW is still used, these days primarily by amateur radio operators (hams). Strictly, on-off keying of a carrier should be known as "Interrupted Continuous Wave" or ICW.

Radio teletypes usually operate on short-wave (HF) and are much loved by the military because they create written information without a skilled operator. They send a bit as one of two tones. Groups of five or seven bits become a character printed by a teletype. From about 1925 to 1975, radio teletype was how most commercial messages were sent to less developed countries. These are still used by the military and weather services.

Aircraft use a 1200 Baud radioteletype service over VHF to send their ID, altitude and position, and get gate and connecting-flight data. Microwave dishes on satellites, telephone exchanges and TV stations usually use quadrature amplitude modulation (QAM). QAM sends data by changing both the phase and the amplitude of the radio signal. Engineers like QAM because it packs the most bits into a radio signal. Usually the bits are sent in "frames" that repeat. A special bit pattern is used to locate the beginning of a frame.

Systems that need reliability, or that share their frequency with other services, may use "corrected orthogonal frequency-division multiplexing" or COFDM. COFDM breaks a digital signal into as many as several hundred slower subchannels. The digital signal is often sent as QAM on the subchannels. Modern COFDM systems use a small computer to make and decode the signal with digital signal processing, which is more flexible and far less expensive than older systems that implemented separate electronic channels. COFDM resists fading and ghosting because the narrow-channel QAM signals can be sent slowly. An adaptive system, or one that sends error-correction codes can also resist interference, because most interference can affect only a few of the QAM channels. COFDM is used for WiFi, some cell phones, Digital Radio Mondiale, Eureka 147, and many other local area network, digital TV and radio standards.

Heating

Radio-frequency energy generated for heating of objects is generally not intended to radiate outside of the generating equipment, to prevent interference with other radio signals. Microwave ovens use intense radio waves to heat food. (Note: It is a common misconception that the radio waves are tuned to the resonant frequency of water molecules. The microwave frequencies used are actually about a factor of ten below the resonant frequency.) Diathermy equipment is used in surgery for sealing of blood vessels. Induction furnaces are used for melting metal for casting.

Mechanical force

Tractor beams can use radio waves which exert small electrostatic and magnetic forces. These are enough to perform station-keeping in microgravity environments. Conceptually, spacecraft propulsion: Radiation pressure from intense radio waves has been proposed as a propulsion method for an interstellar probe called Starwisp. Since the waves are long, the probe could be a very light metal mesh, and thus achieve higher accelerations than if it were a solar sail.

Amateur radio service

Amateur radio is a hobby in which enthusiasts purchase or build their own equipment and use radio for their own enjoyment. They may also provide an emergency and public-service radio service. This has been of great use, saving lives in many instances. Radio amateurs are licensed to use frequencies in a large number of narrow bands throughout the radio spectrum. They use all forms of encoding, including obsolete and experimental ones. Several forms of radio were pioneered by radio amateurs and later became commercially important including FM, single-sideband (SSB), AM, digital packet radio and satellite repeaters. Some amateur frequencies may be disrupted by power-line internet service.

Unlicensed radio services

Personal radio services such as Citizens' Band Radio, Family Radio Service, Multi-Use Radio Service and others exist in North America to provide simple, (usually) short range communication for individuals and small groups, without the overhead of licensing. Similar services exist in other parts of the world. These radio services involve the use of handheld or mobile radios better known as "walkie-talkies".

Radio control (RC)

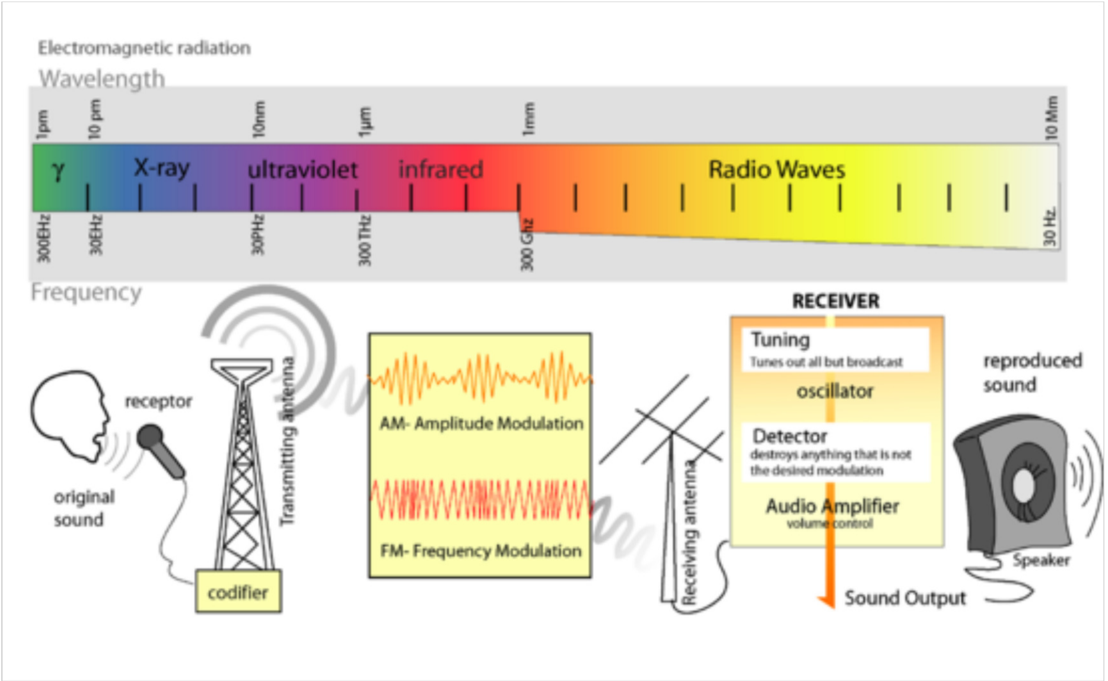
Radio remote control use of radio waves to transmit control data to a remote object as in some early forms of guided missile, some early TV remotes and a range of model boats, cars and airplanes. Large industrial remote-controlled equipment such as cranes and switching locomotives now usually use digital radio techniques to ensure safety and reliability.

In Madison Square Garden, at the Electrical Exhibition of 1898, Nikola Tesla successfully demonstrated a radio-controlled boat. He was awarded U.S. patent No. 613,809 for a "Method of and Apparatus for Controlling Mechanism of Moving Vessels or Vehicles."

The electromagnetic spectrum

Radio waves are a form of electromagnetic radiation, created whenever a charged object (in normal radio transmission, an electron) accelerates with a frequency that lies in the radio frequency (RF) portion of the electromagnetic spectrum. In radio, this acceleration is caused by an alternating current in an antenna. Radio frequencies occupy the range from a few tens of hertz to three hundred gigahertz, although commercially important uses of radio use only a small part of this spectrum.

Other types of electromagnetic radiation, with frequencies above the RF range, are microwave, infrared, visible light, ultraviolet, X-rays and gamma rays. Since the energy of an individual photon of radio frequency is too low to remove an electron from an atom, radio waves are classified as non-ionizing radiation.



Electromagnetic spectrum and diagram of radio transmission of an audio signal. **NB** The colours used in this diagram of the electromagnetic spectrum are for decoration only. They do not correspond to the wavelengths and frequencies indicated on the scale.

Other

Energy autarkic radio technology consists of a small radio transmitter powered by environmental energy (push of a button, temperature differences, light, vibrations, etc.). A number of schemes have been proposed for Wireless energy transfer. Various plans included transmitting power using microwaves, and the technique has been demonstrated. (See Microwave power transmission). These schemes include, for example, solar power stations in orbit beaming energy down to terrestrial users.

Retrieved from " http://en.wikipedia.org/wiki/Radio"

Rail transport

2008/9 Schools Wikipedia Selection. Related subjects: Railway transport


German InterCityExpress

Rail transport is the conveyance of passengers and goods by means of wheeled vehicles specially designed to run along **railways** or **railroads**. Rail transport is part of the logistics chain, which facilitates international trade and economic growth in most countries.

Typical railway tracks consist of two parallel rails, normally made of steel, secured to cross beams, termed *sleepers* (UK and Australia) or *ties* (US). The sleepers maintain a constant distance between the two rails; a measurement known as the " gauge" of the track. To maintain the alignment of the track it is either laid on a bed of ballast or else secured to a solid concrete foundation. The whole is referred to as permanent way (UK and Australia usage) or right-of-way (North American usage).


BNSF Railway freight service in the United States

Railway rolling stock, which is fitted with metal wheels, moves with low frictional resistance when compared to road vehicles. On the other hand, locomotives and powered cars normally rely on the point of contact of the wheel with the rail for traction and adhesion (the part of the transmitted axle load that makes the wheel "adhere" to the smooth rail). While this is usually sufficient under normal dry rail conditions, adhesion can be reduced or even lost through the presence of unwanted material on the rail surface, such as moisture, grease, ice, or dead leaves.

General


Railway tracks running through Stanhope railway station in North East England, UK

Rail transport is an energy-efficient and capital-intensive means of mechanised land transport and is a component of logistics. Along with various engineered components, rails constitute a large part of the permanent way. They provide smooth and hard surfaces on which the wheels of the train can roll with a minimum of friction. As an example, a typical modern wagon can hold up to 125 tons of freight on two four-wheel bogies/trucks (100 tons in UK). The contact area between each wheel and the rail is tiny, a strip no more than a few millimetres wide, which minimizes friction. In addition, the track distributes the weight of the train evenly, allowing significantly greater loads per axle / wheel than in road transport, leading to less wear and tear on the permanent way. This can save energy compared with other forms of transportation, such as road transport, which depends on the friction between rubber tires and the road. Trains also have a small frontal area in relation to the load they are carrying, which cuts down on forward air resistance and thus energy usage, although this does not necessarily reduce the effects of side winds.

Due to these various benefits, rail transport is a major form of public transport in many countries. In Asia, for example, many millions use trains as regular transport in India, China, South Korea and Japan. It is also widespread in European countries. By comparison, intercity rail transport in the United States is relatively scarce outside the Northeast Corridor, although a number of major U.S. cities have heavily-used, local rail-based passenger transport systems or light rail or commuter rail operations.

A railway ticket issued in the United Kingdom

The vehicles travelling on the rails, collectively known as *rolling stock*, are arranged in a linked series of vehicles called a train, which can include a locomotive if the vehicles are not individually powered. A locomotive (or "engine") is a powered vehicle used to haul a train of unpowered vehicles. In the USA, individual unpowered vehicles are known generically as *cars*. These may be passenger carrying or used for freight purposes. For passenger-carrying vehicles, the term *carriage or coach* is used, while a freight-carrying vehicle is known as a *freight car* in the United States and a *wagon* or *truck* in Great Britain. An individually-powered passenger vehicle is known as a *railcar* or a *power car*; when one or more as these are coupled to one or more unpowered *trailer cars* as an inseparable unit, this is called a *railcar set* or *multiple unit*.

History

Stone rails

The earliest evidence of a railway found thus far was the 6-kilometre (3.7 mi) Diolkos wagonway, which transported boats across the Corinth isthmus in Greece during the 6th century BC. Trucks pushed by slaves ran in grooves in limestone, which provided the track element, preventing the wagons from leaving the intended route. The Diolkos ran for over 1300 years, until 900 AD. The first horse-drawn wagonways also appeared in ancient Greece, with others to be found on Malta and various parts of the Roman Empire, using cut-stone tracks. An example of stone track still exists on Dartmoor, England, where the Haytor Granite Tramway was built in 1820 using grooved granite blocks.

Trackwork including a point on the Haytor Granite Tramway

Wooden rails

Railways began reappearing in Europe after the Dark Ages following the collapse of the Roman Empire. The earliest known record of a railway in Europe from this period is a stained-glass window in the Minster of Freiburg im Breisgau dating from around 1350. By 1550, narrow gauge railways operating with wooden rails were common in mines in Europe. The first railways in Great Britain (also known as wagonways) were constructed in the early 17th century, mainly for transporting coal from mines to canal wharfs where it could be transferred to a boat for onward shipment. The earliest recorded examples are the Wollaton Wagonway in Nottinghamshire and the Bourtreehill - Broomlands Wagonway in Irvine, Ayrshire. Other examples can be found in Broseley in Shropshire, where wooden rails and flanged wheels were utilised, as on a modern railway. However, the rails were prone to wear out under the pressure, and had to be replaced regularly.

Iron plate rail

In 1768, the Coalbrookdale Iron Works laid cast iron plates on top of the wooden rails, providing a more durable load-bearing surface. These were later used by Benjamin Outram at his foundry in Ripley, Derbyshire, the first time standardised components were produced. It was these that led to the name "platelayer" for workers on the permanent way. The advantage was that a considerable variation in wheel spacing (gauge) could be accommodated. However, wheels would bind against the upright part of the plate, and mud and stones would accumulate. On the Little Eaton Gangway in 1799, where Outram used passing loops on the single track, moveable plates were provided, called "pointers", which became shortened to "points".

Edge rail

Lengths of "fishbelly" rail on stone support blocks

From the late 18th century, iron "edge rails" began to appear. The British civil engineer William Jessop designed smooth iron edge rails, which were used in conjunction with flanged iron wheels, introducing them on a route between Loughborough and Nanpantan, Leicestershire, as an adjunct to the Charnwood Forest Canal, in 1793-4. In 1803, Jessop opened the Surrey Iron Railway in south London, arguably the world's first horse-drawn public railway. Being of cast iron these rails were short, around three feet long, of a "fish-bellied" design. They had a foot at each end by means of which they were fastened to stone blocks in the ground.

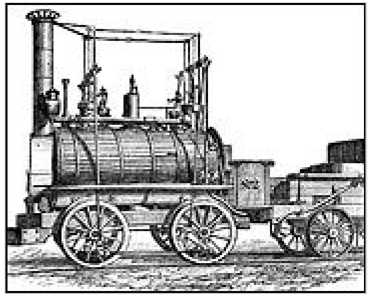
Wrought iron and steel

Cast iron is a brittle material and the short lengths meant that they soon became uneven. However, developments in the process of hot rolling iron meant that longer length rails could be produced. In 1805, the first wrought iron rails were produced at Bedlington Ironworks near Durham. The first steel rails were produced by Robert Forester Mushet and laid at Derby station in 1857. Modern railways still use steel rails, but they are typically welded together to form lengths of *continuous welded rail* which removes the additional wear and tear on rolling stock caused by the tiny differences in rail surface height at the joint between adjacent rail sections.

Motive power

Steam locomotives

The first locomotive to haul a train of wagons on rails was designed by Cornish engineer Richard Trevithick, and was demonstrated in 1804 on a plateway at Merthyr Tydfil, South Wales. Although the locomotive successfully hauled the train, the rail design was not a success, partly because its weight broke a number of the brittle cast-iron plates. Despite this setback, another area of South Wales pioneered rail operations, when, in 1806, a horse-drawn railway was built between Swansea and Mumbles: the Swansea-Mumbles railway started carrying fare-paying passengers in 1807 – the first in the world to do so.



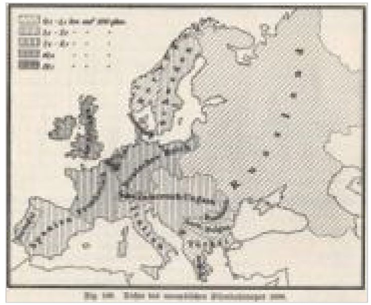
Blücher, an early railway locomotive built in 1814 by George Stephenson

In 1811, John Blenkinsop designed the first successful and practical railway locomotive. He patented a system of moving coals by a rack railway worked by a steam locomotive (patent no. 3431), and a line was built connecting the Middleton Colliery to Leeds. The locomotive (*The Salamanca*) was built in 1812 by Matthew Murray of Fenton, Murray and Wood. The Middleton Railway was the first railway to successfully use steam locomotives on a commercial basis. It was also the first railway in Great Britain to be built under the terms laid out in an Act of Parliament. Blenkinsop's engine had double-acting cylinders and, unlike the Trevithick pattern, no flywheel. Due to previous experience with broken rails, the locomotive was made very light and this brought concerns about insufficient adhesion, so instead of driving the wheels directly, the cylinders drove a cogwheel through spur gears, the cogwheel providing traction by engaging with a rack cast into the side of the rail.

In Scotland, the Kilmarnock and Troon Railway was the first railway constructed, and was authorised by Act of Parliament in 1808. The civil engineer leading the project was William Jessop, and its 1811 construction meant that it was the first railway in Scotland to use a steam locomotive, while it was the only line in Scotland for 14 years. Its representation appeared in the Coat of Arms of the Burgh of Troon. The line was intended to carry coal for the Duke of Portland; and ran services between Kilmarnock and Troon Harbour. The line began life as a 9.5 mile (16 km), double track 4 ft 0 in (1,219 mm) gauge, horse-drawn waggonway. It was built using cast iron plate rails with an inner flange. A George Stephenson-built locomotive, his second one from Killingworth Colliery, was tried on the main line in 1817, but the weight of the engine broke the cast iron plate rails. It worked better when wooden rails were used, and the locomotive remained in use until 1848.

Magic lantern image of Lahore Railway Station, Lahore circa 1895

The Stockton and Darlington Railway opened in northern England in 1825 to be followed five years later by the Liverpool and Manchester Railway, considered to be the world's first "Inter City" line. The rail gauge (the distance between the two rails of the track) was used for the early wagonways, and had been adopted for the Stockton and Darlington Railway. The 4 ft 8½ in (1,435 mm) width became known as the international " standard gauge", used by about 60 percent of the world's railways. The Liverpool and Manchester Railway, on the other hand, proved the viability of rail transport when, after organising the Rainhill Trials of 1829, Stephenson's *Rocket* successfully hauled a load of 13 tons at an average speed of 12 miles per hour. The company took the step of working its trains from its opening entirely by steam traction. Railways then soon spread throughout the United Kingdom and the world, and became the dominant means of land transport for nearly a century, until the invention of aircraft and automobiles, which prompted a gradual decline in railways.



Density of the railway net in Europe 1896

The first railroad in the United States may have been a gravity railroad in Lewiston, New York in 1764. The 1810 Leiper Railroad in Pennsylvania was intended as the first permanent railroad, and the 1826 Granite Railway in Massachusetts was the first commercial railroad to evolve through continuous operations into a common carrier. The Baltimore and Ohio, opened in 1830, was the first to evolve into a major system. In 1867, the first elevated railroad was built in New York. In 1869, the symbolically important transcontinental railroad was completed in the United States with the driving of a golden spike at Promontory, Utah. The development of the railroad in the United States helped reduce transportation time and cost, which allowed migration towards the west. Railroads increased the accessibility of goods to consumers, thus allowing individuals and capital to flow westward. Railroads created national markets characterized by the 'law of one price' by lowering difference in price charged for commodity between suppliers and demanders. Railroads increased social savings, and were the largest contributors of any innovation before 1900.

The South American experience regarding railways was first achieved in 1854, when a line was laid between the Chilean towns of Caldera and Copiapo. However, the first concerted trans-Andine attempt between Argentina and Chile did not occur until the 1870s, due to the financial risks involved in such a project. It was not until 1887 that the Argentinians began to construct their part of the enterprise, with the Chileans beginning construction in 1889, though by 1893, work had ceased due to financial constraints. In 1896, the Transandine Railway Company was created in London to purchase the existing railways and construct a continuous line between Argentina and Chile that would improve transport and communication links in South America. This was finally completed in 1908, when the Argentine and Chilean stretches of track were joined.

Dieselisation

Two SD70M diesel locomotives of the Union Pacific refuelling at Dunsmuir, California

Dieselisation was the replacement of the steam locomotive with the diesel-electric locomotive (often referred to as a " diesel locomotive"), a process which began in the 1930s and is now substantially complete worldwide.

Dieselisation took place largely because of the reduction in operating costs it allowed. Steam locomotives require large pools of labour to clean, load, maintain and run. They also require extensive service, coaling and watering facilities. Diesel locomotives require significantly less time and labour to operate and maintain.

After World War II, dramatically increased labour costs in the Western World made steam an increasingly costly form of motive power. At the same time, the war had forced improvements in internal combustion engine technology that made diesel locomotives cheaper and more powerful. The post war world also re-aligned the business and financial markets, as did world geo-politics as in the Cold War (1947-1953).

Electrification

Robert Davidson started to experiment with an electrical railway car in Scotland in 1838. By 1839 he had completed and presented a 4.8 m long carriage that weighed six tons, including batteries. It reached a maximum speed of 6.4 kilometres per hour.

Postcard showing electric streetcars in Richmond, Virginia, where Frank J. Sprague successfully demonstrated his new system on the hills in 1888

Magnus Volk opened his electric railway in Brighton in 1883.

The use of overhead wires conducting electricity, invented by Granville T. Woods in 1888, among several other improvements, led to the development of electrified railways, the first of which in the United States was operated at Coney Island in 1892. Richmond, Virginia had the first successful electrically-powered trolley system in the United States. Designed by electric power pioneer Frank J. Sprague, the trolley system opened its first line in January, 1888. Richmond's hills, long a transportation obstacle, were considered an ideal proving ground. The new technology soon replaced horse-powered streetcars.

Sweden got the perhaps first fully electrified developed railway that efficiently transported commuters as well as goods, in 1895. At the time it ran from close to central Stockholm to Rimbo, located in the countryside Roslagen. It is still in use to commuters today but runs only about a third of its biggest extent, much due to it not using the standard gauge but 3ft (891mm).

Japanese Shinkansen train passing Mount Fuji

In the USSR the phenomenon of children's railways was developed in the 1930s (the world's first one was opened on 24 July 1935). Fully operated by children, they were extracurricular educational institutions, where teenagers learned railway professions. A lot of them are functioning in post-Soviet states and Eastern European countries.

Many countries since the 1960s have adopted high-speed railways. On 3 April 2007, the French TGV set a new train speed record. The train, with a modified engine and wheels, reached 574.8 km/h (357.2 mph). The record attempt took place on the new LGV Est line between Paris and Strasbourg using a specially equipped TGV Duplex train. The overhead lines had also been modified for the attempt to carry 31,000 V rather than the line's normal 25,000 V. On 24 August 2005, the Qingzang railway became the highest railway line in the world, when track was laid through the Tanggula Mountain Pass at 5,072 meters (16,640 ft) above sea level in the Tanggula Mountains, Tibet.

Operations

A railway can be broken down into two major components. Firstly, there are the items which "move", also referred to as the rolling stock, which include locomotives, passenger carrying vehicles (or coaches), freight carrying vehicles (or goods wagons). Secondly are the "fixed" components, usually referred to as the railway's infrastructure, including the permanent way and ancillary buildings that are necessary for a railway to function.

Rolling stock

A locomotive is the vehicle that provides the motive power for a train. A locomotive has no payload capacity of its own, and its sole purpose is to move the train along the tracks. Traditionally, locomotives pull trains from the front.

A railroad car is a vehicle used for the haulage of either passengers or freight. Most cars carry a "revenue" load, although "non-revenue" cars exist for the railroad's own use, such as for maintenance-of-way purposes.

Two British Rail Class 143 DMUs at Cardiff Queen Street station in the United Kingdom

Signalling

Railway signalling is a system used to control railway traffic safely to prevent trains from colliding. Being guided by fixed rails, trains are uniquely susceptible to collision since they frequently operate at speeds that do not enable them to stop quickly or, in some cases, within the driver's sighting distance.

Most forms of train control involve movement authority being passed from those responsible for each section of a rail network (e.g., a signalman or stationmaster) to the train crew. The set of rules and the physical equipment used to accomplish this control determine what is known as the *method of working* (UK), *method of operation* (US) or *safeworking* (Aus.). Not all methods require the use of signals, and some systems are specific to single track railways. The signalling process is traditionally carried out in a signal box or interlocking tower, a small building that houses the lever frames required for the signalman to operate switches and signal equipment. These are placed at various intervals along the route of a railway, controlling specified sections of track. More recent technological developments have made such operational doctrine superfluous, with the centralization of signalling operations to regional control rooms. This has been facilitated by the increased use of computers, allowing vast sections of track to be monitored from a single location.

GWR semaphore-type signal

Right of way

Railway tracks are laid upon land owned or leased by the railway. Owing to the requirements for large radius turns and modest grades, rails will often be laid in circuitous routes. Public carrier railways are typically granted limited rights of eminent domain (UK:compulsory purchase). In many cases in the 19th century, railways were given additional incentives in the form of grants of public land. Route length and grade requirements can be reduced by the use of alternating earthen cut and fill, bridges, and tunnels, all of which can greatly increase the capital expenditures required to develop a right of way, while significantly reducing operating costs and allowing higher speeds on longer radius curves. In densely urbanized areas such as Manhattan, railways are sometimes laid out in tunnels to minimize the effects on existing properties (see condemnation).

Safety and railway disasters

Trains can travel at very high speed; however, they are heavy, are unable to deviate from the track and require a great distance to stop. Although rail transport is considered one of the safest forms of travel, there are many possibilities for accidents to take place. These can vary from the minor derailment (jumping the track), a head-on collision with another train and collision with an automobile or other vehicle at a level crossing/grade crossing. Level crossing collisions are relatively common in the United States where there are several thousand each year killing about 500 people (the comparable figures in the United Kingdom are 30 and 12 collisions and casualties, respectively). For information regarding major accidents, see List of rail accidents.


Train wreck, 1907, in Canaan, New Hampshire

The most important safety measures are railway signalling and gates at level/grade crossings. Train whistles warn of the presence of a train, while trackside signals maintain the distances between trains. In the United Kingdom, vandalism or negligence is thought responsible for about half of rail accidents. Railway lines are zoned or divided into blocks guarded by combinations of block signals, operating rules, and automatic-control devices so that one train, at most, may be in a block at any time.

Compared with road travel, railways remain relatively safe. Annual death rates on roads are over 40,000 in the United States, about 3,000 in the United Kingdom and 900 in Australia, compared with 1,000 rail-related fatalities in the United States,under 20 in the UK and 10 in Australia. (These figures do not account for differences in passenger-miles traveled by mode; see e.g. Transportation safety in the United States.)

Trackage

A typical railway/railroad track consists of two parallel steel (or in older networks, iron) rails, generally anchored perpendicular to beams, termed sleepers or ties, of timber, concrete, or steel to maintain a consistent distance apart, or gauge. The rails and perpendicular beams are usually then placed on a foundation made of concrete or compressed earth and gravel in a bed of ballast to prevent the track from buckling (bending out of its original configuration) as the ground settles over time under the weight of the vehicles passing above. The vehicles traveling on the rails are arranged in a train; a series of individual powered or unpowered linked vehicles, displaying markers. These vehicles (referred to, in general, as *cars, carriages or wagons*) move with much less friction than do vehicles riding on rubber tires on a paved road, and the locomotive that pulls the train tends to use energy far more efficiently as a result.


Concrete ties (sleepers)

Trackage, consisting of sleepers/ties and rails, may be prefabricated or assembled in place. Rails may be composed of segments welded or bolted, and may be of a length comparable to that of a railcar or two or may be many hundreds of feet long.

The surface of the ballast is sloped around curves to reduce lateral forces. This reduces the forces tending to displace the track, reduces the tendency to overturn at high speed, and makes for a more comfortable ride for standing cattle and standing or seated passengers. This will be optimal at only one particular speed, however.


Trestle bridge

Track components

Railways are highly complex feats of engineering, with many hours of planning and forethought required for a successful outcome. The first component of a railway is the route, which is planned to provide the least resistance in terms of gradient and engineering works. As such, the track bed is heavily engineered to provide, where possible, a level surface. As such, embankments are constructed to support the track and to provide a compromise in terms of the route's average elevation. With this in mind, sundry structures such as bridges and viaducts are constructed in an attempt to maintain the railway's elevation, and gradients are kept within manageable constraints. Where such structures are not always justified, such as in hilly terrain where routes may require long detours to avoid such features, a cutting or tunnel is dug or bored through the obstacle. Once the sundry engineering works are completed, a bed of stone (ballast) is laid over the compacted track bed to enhance drainage around the ties and evenly distribute pressure over a wider area, locking the track-work in place. Crushed stone is firmly tamped to prevent further settling and to lock the stones. Minor water courses are channeled through pipes (culverts) before the grade is raised


Bolted rail connection and tie-down. Also known as a fishplate.

The base of the trackage consists of treated wood or concrete " ties", also known as "sleepers". These ensure the proper distance between the rails (known as "gauge") and anchor the rail structure to the road bed through the use of plates. These are attached to the top of the ties to provide a secure housing for the rails. After placement of the rail atop the plate, spikes are driven through holes in the plate and into the tie where they are held by friction. The top of the spike has a head that clamps the rail. As an alternative, lag bolts can be used to retain the clamps, which is preferred since screws are less likely to loosen. The space between and surrounding the ties is filled with additional ballast to stabilize the rail assembly.

Points (Turnouts or Switches)

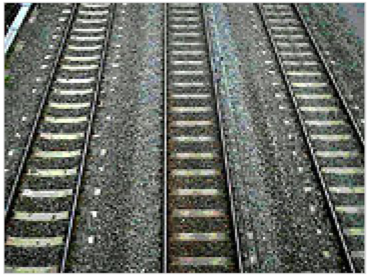

Railway turnouts

Points (UK) or switches (US), technically known as turnouts, are the means of directing a train onto a diverging section of track, for example, a siding, a branch line, or a parallel running line. Laid similar to normal track, a point typically consists of a frog (common crossing), check rails and two switch rails. The switch rails may be moved left or right, under the control of the signalling system, to determine which path the train will follow.

Maintenance

Spikes in wooden ties can loosen over time, while split and rotten ties may be individually replaced with a concrete substitute. Should the rails settle due to soil subsidence, they can be lifted by specialized machinery and additional ballast tamped down to form a level bed. Periodically, ballast must be removed and replaced with clean ballast to ensure adequate drainage, especially if wooden ties are used. Culverts and other passages for water must be kept clear lest water is impounded by the trackbed, causing landslips. Where trackbeds are placed along rivers, additional protection is usually placed to prevent erosion during times of high water, while bridges are another important item requiring inspection and maintenance.

Terminology


Rail tracks

In the United Kingdom and most other Commonwealth of Nations countries, the term *railway* is used in preference to the United States term, *railroad.* In Canadian speech, *railway* and *railroad* are interchangeable, although in law *railway* is the usual term. *Railroad* was used in the United Kingdom concurrently with *railway* until the 1850s when *railway* became the established term. Several American companies have *railway* in their names instead of *railroad*, the BNSF Railway being the pre-eminent modern example.

In the United Kingdom, the term *railway* often refers to the whole organization of tracks, trains, stations, signalling, timetables and the operating companies that collectively make up a coordinated railway system, while *permanent way* or *p/way* refers to the tracks alone; however this terminology is generally not commonplace outside of the railway industry or those who take a keen interest in it.

Subways, metros, elevated lines, trolley lines, and undergrounds are all specialized railways.

Rail transport by country

Of 236 countries and dependencies, 143 have rail transport (including several with very little), of which about 90 have passenger services.

Steam engine

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

A scale model Allchin traction engine
– an example of a self-propelled steam engine

A **steam engine** is a heat engine that performs mechanical work using steam as its working fluid.

Steam engines have a long history, going back almost two thousand years. Early devices were not practical power producers, but more advanced designs become a major source of mechanical power during the industrial revolution. Modern steam turbines generate about half of the electric power in the world.

Many steam engines are external combustion engines, although other sources of heat such as solar power, nuclear power or geothermal energy are often used. The heat cycle used is known as the Rankine cycle.

In general usage, the term 'steam engine' can refer to integrated steam plants such as railway steam locomotives and portable engines, or may refer to the motor unit alone, as in the beam engine and stationary steam engine. Specialized devices such as steam hammers and steam pile drivers are dependent on steam supplied from a separate, often remotely-located boiler.

Applications

'Preserved' (but incomplete) portable engine, Tenterfield, NSW – an example of a mobile steam engine

Since the early 18th century steam power has been set to a variety of practical uses. At first it was applied to reciprocating pumps, but from the 1780s rotative engines (i.e. those converting reciprocating motion into rotary motion) began to appear, driving factory machinery. At the turn of the 19th century, steam-powered transport on both sea and land began to make its appearance becoming ever more predominant as the century progressed.

Steam engines can be said to have been the moving force behind the Industrial Revolution and saw widespread commercial use driving machinery in factories and mills, powering pumping stations and transport appliances such as locomotives, steam ship engines and road vehicles. Their use in agriculture led to an increase in the land available for cultivation.

Very low power engines are used to power models and speciality applications such as the steam clock.

The presence of several phases between heat source and power delivery has meant that it has always been difficult to obtain a power-to-weight ratio anywhere near that obtainable from internal combustion engines; notably this has made steam aircraft extremely rare. Similar considerations have meant that for small and medium-scale applications steam has been largely superseded by internal combustion engines or electric motors, which has given the steam engine an out-dated image. However it is important to remember that the power supplied to the electric grid is predominantly generated using steam turbine plant, so that indirectly the world's industry is still dependent on steam power. Recent concerns about fuel sources and pollution have incited a renewed interest in steam both as a component of cogeneration processes and as a prime mover. This is becoming known as the Advanced Steam movement.

Steam engines can be classified by their application:

Stationary applications

Stationary steam engines can be classified into two main types:

1. Winding engines, rolling mill engines, steam donkeys, marine engines, and similar applications which need to frequently stop and reverse.
2. Engines providing power, which stop rarely and do not need to reverse. These include engines used in thermal power stations and those that were used in pumping stations, mills, factories and to power cable railways and cable tramways before the widespread use of electric power.

The steam donkey is technically a stationary engine but is mounted on skids to be semi-portable. It is designed for logging use and can drag itself to a new location. Having secured the winch cable to a sturdy tree at the desired destination, the machine will move towards the anchor point as the cable is winched in.

A portable engine is a stationary engine mounted on wheels so that it may be towed to a work-site by horses or a traction engine, rather than being fixed in a single location.

Transport applications

Steam engines have been used to power a wide array of transport appliances:

- Marine: Steamboat, Steamship
- Rail: Steam locomotive, Fireless locomotive
- Agriculture: Traction engine, Steam tractor
- Road: Steam wagon, Steam bus, Steam tricycle, Steam car
- Construction: Steam roller, Steam shovel
- Military: Steam tank (tracked), Steam tank (wheeled)
- Space: Steam rocket

In many mobile applications internal combustion engines are more frequently used due to their higher power-to-weight ratio, steam engines are used when higher efficiency is needed and weight is less of an issue.

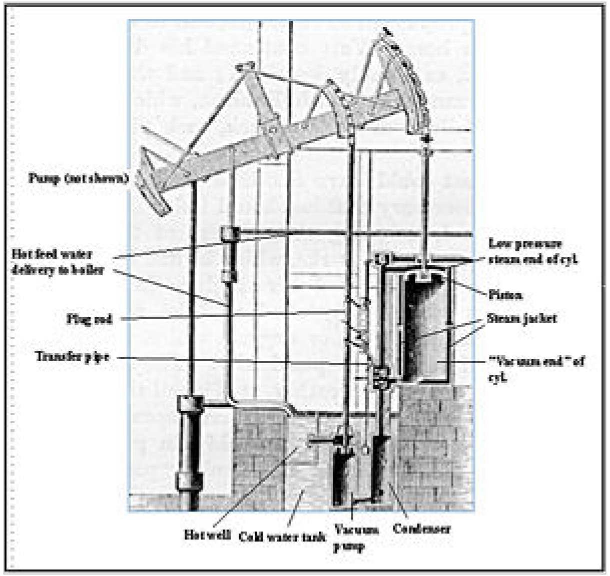
History

The history of the steam engine stretches back as far as the first century AD; the first recorded steam engine being the aeolipile described by Hero of Alexandria. In the following centuries, the few engines known about were essentially experimental devices used by inventors to demonstrate the properties of steam, such as the steam turbines developed by Taqi al-Din in 1551 and Giovanni Branca in 1629.


Aeolipile

The first practical steam-powered 'engine' was a water pump, developed in 1698 by Thomas Savery. It proved only to have a limited lift height and was prone to boiler explosions, but it still received some use for mines and pumping stations.

The first commercially-successful engine did not appear until 1712. Incorporating technologies discovered by Savery and Denis Papin, the atmospheric engine, invented by Thomas Newcomen, paved the way for the Industrial Revolution. Newcomen's engine was relatively inefficient, and in most cases was only used for pumping water. It was mainly employed for draining mine workings at depths hitherto impossible, but also for providing a reusable water supply for driving waterwheels at factories sited away from a suitable 'head'.



Early Watt pumping engine.

The next major step occurred when James Watt developed an improved version of Newcomen's engine. Watt's engine used 75% less coal than Newcomen's, and was hence much cheaper to run. Watt proceeded to develop his engine further, modifying it to provide a rotary motion suitable for driving factory machinery. This enabled factories to be sited away from rivers, and further accelerated the pace of the Industrial Revolution.

Around 1800, Richard Trevithick introduced engines using high-pressure steam. These were much more powerful than previous engines and could be made small enough for transport applications. Thereafter, technological developments and improvements in manufacturing techniques (partly brought about by the adoption of the steam engine as a power source) resulted in the design of more efficient engines that could be smaller, faster, or more powerful, depending on the intended application.

Steam engines remained the dominant source of power well into the 20th century, when advances in the design of electric motors and internal combustion engines gradually resulted in the vast majority of reciprocating steam engines being replaced in commercial usage, and the ascendency of steam turbines in power generation.

See also

The history of steam engine development is a vast subject. The following articles cover aspects of steam engine development in greater detail:

- Timeline of steam power – *overview*
- History of the steam engine – *general history, concentrating on reciprocating engines*
- Steam turbine – *the parallel development of turbine-type engines*
- Steam power during the Industrial Revolution
- History of steam road vehicles

Basic operation of a simple steam engine

- Heat is obtained from fuel burnt in a closed firebox
- The heat is transferred to the water in a pressurised boiler, ultimately boiling the water and transforming it into saturated steam. Steam in its saturated state is always produced at the temperature of the boiling water, which in turn depends on the steam pressure on the water surface within the boiler.
- The steam is transferred to the motor unit which uses it to push on pistons to power machinery.
- The used, cooler, lower pressure steam is dumped to the environment.

Components of steam engines

There are two fundamental components of a steam engine: the boiler or steam generator, and the motor unit, itself often referred to as a "steam engine". The two components can either be integrated into a single unit or can be placed at a distance from each other, in a variety of configurations.

Other components are often present; pumps to supply water to the boiler during operation, condensers to recirculate the water and recover the latent heat of vaporisation, and superheaters to raise the temperature of the steam above its saturated vapour point, and various mechanisms to increase the draft for fireboxes.

Cold sink

As with all heat engines, a considerable quantity of waste heat is produced at relatively low temperature. This must be disposed of.

The simplest cold sink is simply to vent the steam to the environment. This is often used on Steam locomotives, but is quite inefficient. Steam locomotive condensing apparatus can be employed to improve efficiency.

Steam turbines in power stations often use cooling towers which are essentially one form of condenser.

Sometimes the 'waste heat' is useful in and of itself, and in those cases very high overall efficiency can be obtained; for example combined heat and power uses the waste heat for district heating.

Boilers

Boilers are pressure vessels that contain water to be boiled, and some kind of mechanism for transferring the heat to the water so as to boil it.

The two most common methods of transferring heat to the water according are:

1. water tube boiler - water is contained in or run through one or several tubes surrounded by hot gases
2. firetube boiler - the water partially fills a vessel below or inside of which is a combustion chamber or furnace and fire tubes through which the hot gases flow

Once turned to steam, some boilers use superheating to raise the temperature of the steam further. This allows for greater efficiency.

Motor units

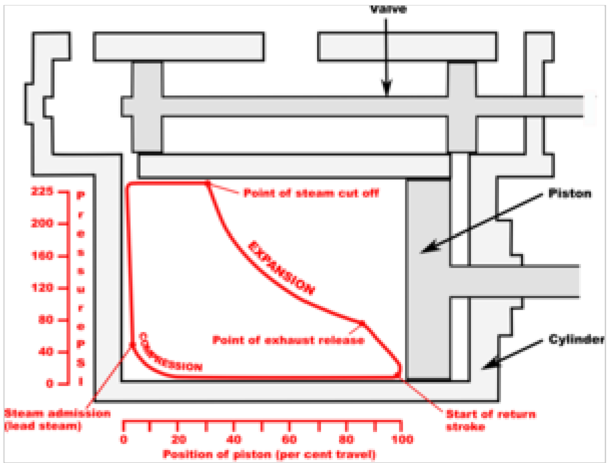
A motor unit takes a supply of steam at high pressure and temperature and gives out a supply of steam at lower pressure and temperature, using as much of the difference in steam energy as possible to do mechanical work.

A motor unit is often called 'steam engine' in its own right. They will also operate on compressed air or other gas.

Simple expansion

This means that a charge of steam works only once in the cylinder. It is then exhausted directly into the atmosphere or into a condenser, but remaining heat can be recuperated if needed to heat a living space, or to provide warm feedwater for the boiler.

In most reciprocating piston engines the steam reverses its direction of flow at each stroke (counterflow), entering and exhausting from the cylinder by the same port. The complete engine cycle occupies one rotation of the crank and two piston strokes; the cycle also comprises four *events — admission, expansion, exhaust, compression*. These events are controlled by valves often working inside a *steam chest* adjacent to the cylinder; the valves distribute the steam by opening and closing steam *ports* communicating with the cylinder end(s) and are driven by valve gear, of which there are many types.



Schematic Indicator diagram showing the four events in a double piston stroke

The simplest valve gears give events of fixed length during the engine cycle and often make the engine rotate in only one direction. Most however have a reversing mechanism which additionally can provide means for saving steam as speed and momentum are gained by gradually "shortening the cutoff" or rather, shortening the admission event; this in turn proportionately lengthens the expansion period. However, as one and the same valve usually controls both steam flows, a short cutoff at admission adversely affects the exhaust and compression periods which should ideally always be kept fairly constant; if the exhaust event is too brief, the totality of the exhaust steam cannot evacuate the cylinder, choking it and giving excessive compression (*"kick back"*).

In the 1840s and 50s there were attempts to overcome this problem by means of various patent valve gears with separate variable cutoff valves riding on the back of the main slide valve; the latter usually had fixed or limited cutoff. The combined setup gave a fair approximation of the ideal events, at the expense of increased friction and wear, and the mechanism tended to be complicated. The usual compromise solution has been to provide *lap* by lengthening rubbing surfaces of the valve in such a way as to overlap the port on the admission side, with the effect that the exhaust side remains open for a longer period after cut-off on the admission side has occurred. This expedient has since been generally considered satisfactory for most purposes and makes possible the use of the simpler Stephenson, Joy and Walschaerts motions. Corliss, and later, poppet valve gears had separate admission and exhaust valves driven by trip mechanisms or cams profiled so as to give ideal events; most of these gears never succeeded outside of the stationary marketplace due to various other issues including leakage and more delicate mechanisms.

Compression

Before the exhaust phase is quite complete, the exhaust side of the valve closes, shutting a portion of the exhaust steam inside the cylinder. This determines the compression phase where a cushion of steam is formed against which the piston does work whilst its velocity is rapidly decreasing; it moreover obviates the pressure and temperature shock, which would otherwise be caused by the sudden admission of the high pressure steam at the beginning of the following cycle.

Lead

The above effects are further enhanced by providing *lead*: as was later discovered with the internal combustion engine, it has been found advantageous since the late 1830s to advance the admission phase, giving the valve *lead* so that admission occurs a little before the end of the exhaust stroke in order to fill the *clearance volume* comprising the ports and the cylinder ends (not part of the piston-swept volume) before the steam begins to exert effort on the piston.

Compounding engines

As steam expands in a high pressure engine its temperature drops; because no heat is released from the system, this is known as adiabatic expansion and results in steam entering the cylinder at high temperature and leaving at low temperature. This causes a cycle of heating and cooling of the cylinder with every stroke which is a source of inefficiency.

A method to lessen the magnitude of this heating and cooling was invented in 1804 by British engineer Arthur Woolf, who patented his *Woolf high pressure **compound engine*** in 1805. In the compound engine, high pressure steam from the boiler expands in a high pressure (HP) cylinder and then enters one or more subsequent lower pressure (LP) cylinders. The complete expansion of the steam now occurs across multiple cylinders and as less expansion now occurs in each cylinder so less heat is lost by the steam in each. This reduces the magnitude of cylinder heating and cooling, increasing the efficiency of the engine. To derive equal work from lower pressure steam requires a larger cylinder volume as this steam occupies a greater volume. Therefore the bore, and often the stroke, are increased in low pressure cylinders resulting in larger cylinders.

Double expansion (usually known as **compound**) engines expanded the steam in two stages. The pairs may be duplicated or the work of the large LP cylinder can be split with one HP cylinder exhausting into one or the other, giving a 3-cylinder layout where cylinder and piston diameter are about the same making the reciprocating masses easier to balance.

Two-cylinder compounds can be arranged as:

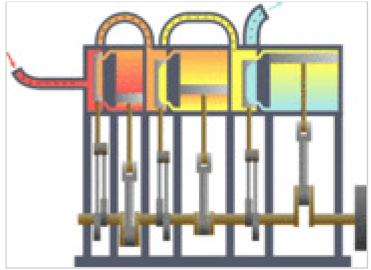
- **Cross compounds** - The cylinders are side by side.
- **Tandem compounds** - The cylinders are end to end, driving a common connecting rod
- **Angle compounds** - The cylinders are arranged in a vee (usually at a 90° angle) and drive a common crank.

With two-cylinder compounds used in railway work, the pistons are connected to the cranks as with a two-cylinder simple at 90° out of phase with each other (*quartered*). When the double expansion group is duplicated, producing a 4-cylinder compound, the individual pistons within the group are usually balanced at 180°, the groups being set at 90° to each other. In one case (the first type of Vauclain compound), the pistons worked in the same phase driving a common crosshead and crank, again set at 90° as for a two-cylinder engine. With the 3-cylinder compound arrangement, the LP cranks were either set at 90° with the HP one at 135° to the other two, or in some cases all three cranks were set at 120°.

The adoption of compounding was common for industrial units, for road engines and almost universal for marine engines after 1880; it was not universally popular in railway locomotives where it was often perceived as complicated. This is partly due to the harsh railway operating environment and limited space afforded by the loading gauge (particularly in Britain, where compounding was never common and not employed after 1930). However although never in the majority it was popular in many other countries

Multiple expansion engines

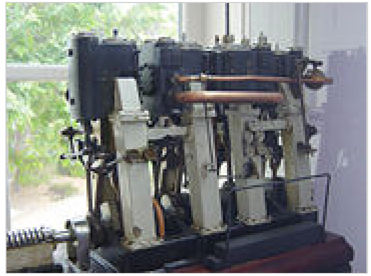

1890s-vintage triple-expansion marine engine that powered the SS *Christopher Columbus*


An animation of a simplified triple-expansion engine. High-pressure steam (red) enters from the boiler and passes through the engine, exhausting as low-pressure steam (blue) to the condenser.

It is a logical extension of the compound engine (described above) to split the expansion into yet more stages to increase efficiency. The result is the **multiple expansion engine**. Such engines use either three or four expansion stages and are known as *triple* and *quadruple expansion engines* respectively. These engines use a series of double-acting cylinders of progressively increasing diameter and/or stroke and hence volume. These cylinders are designed to divide the work into three or four, as appropriate, equal portions for each expansion stage. As with the double expansion engine, where space is at a premium, two smaller cylinders of a large sum volume may be used for the low pressure stage. Multiple expansion engines typically had the cylinders arranged inline, but various other formations were used. In the late 19th century, the Yarrow-Schlick-Tweedy balancing 'system' was used on some marine triple expansion engines. Y-S-T engines divided the low pressure expansion stages between two cylinders, one at each end of the engine. This allowed the crankshaft to be better balanced, resulting in a smoother, faster-responding engine which ran with less vibration. This made the 4-cylinder triple-expansion engine popular with large passenger liners (such as the Olympic class), but was ultimately replaced by the virtually vibration-free turbine (see below).


SS *Ukkopekka* triple expansion steam engine

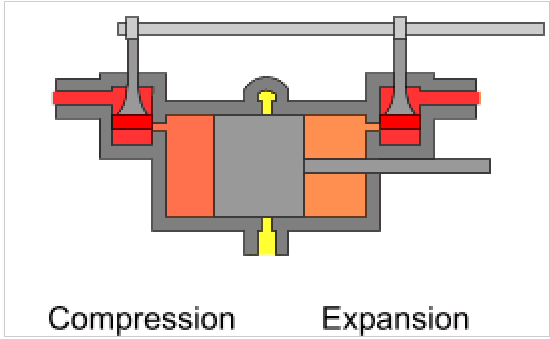
The image to the right shows an animation of a triple expansion engine. The steam travels through the engine from left to right. The valve chest for each of the cylinders is to the left of the corresponding cylinder.


Model of a triple expansion engine

The development of this type of engine was important for its use in steamships as by exhausting to a condenser the water can be reclaimed to feed the boiler, which is unable to use seawater. Land-based steam engines could exhaust much of their steam, as feed water was usually readily available. Prior to and during World War II, the expansion engine dominated marine applications where high vessel speed was not essential. It was however superseded by the British invention steam turbine where speed was required, for instance in warships, such as the pre-dreadnought battleships, and ocean liners. HMS *Dreadnought* of 1905 was the first major warship to replace the proven technology of the reciprocating engine with the then-novel steam turbine.

Uniflow (or unaflow) engine

This is intended to remedy the difficulties arising from the usual counterflow cycle mentioned above which means that at each stroke the port and the cylinder walls will be cooled by the passing exhaust steam, whilst the hotter incoming admission steam will waste some of its energy in restoring working temperature. The aim of the uniflow is to remedy this defect by providing an additional port uncovered by the piston at the end of its half-stroke making the steam flow only in one direction. By this means, thermal efficiency is improved by having a steady temperature gradient along the cylinder bore. The simple-expansion uniflow engine is reported to give efficiency equivalent to that of classic compound systems with the added advantage of superior part-load performance. It is also readily adaptable to high-speed uses and was a common way to drive electricity generators towards the end of the 19th century before the coming of the steam turbine.



Schematic animation of a uniflow steam engine. The poppet valves are controlled by the rotating camshaft at the top. High pressure steam enters, red, and exhausts, yellow.

Uniflow engines have been produced in single-acting, double-acting, simple, and compound versions. Skinner 4-crank 8-cylinder single-acting tandem compound engines power two Great Lakes ships still trading today (2007). These are the *Saint Marys Challenger*, that in 2005 completed 100 years of continuous operation as a powered carrier (the Skinner engine was fitted in 1950) and the car ferry, *S. S. Badger*.

In the early 1950s the Ultimax engine, a 2-crank 4-cylinder arrangement similar to Skinner's, was developed by Abner Doble for the Paxton car project with tandem opposed single-acting cylinders giving effective double-action.

Turbine engines

A **steam turbine** consists of an alternating series of rotating discs mounted on a drive shaft, *rotors*, and static discs fixed to the turbine casing, *stators*. The rotors have a propeller-like arrangement of blades at the outer edge. Steam acts upon these blades, producing rotary motion. The stator consists of a similar, but fixed, series of blades that serve to redirect the steam flow onto the next rotor stage. A steam turbine often exhausts into a condenser that provides a vacuum. The stages of a steam turbine are typically arranged to extract the maximum potential work from a specific velocity and pressure of steam, giving rise to a series of variably sized high and low pressure stages. Turbines rotate at very high speed, therefore are usually connected to reduction gearing to drive another mechanism, such as a ship's propeller, at a lower speed. A turbine rotor is also capable of providing power when rotating in one direction only. Therefore a reversing stage or gearbox is usually required where power is required in the opposite direction.


A rotor of a modern **steam turbine**, used in a power plant

Steam turbines provide direct rotational force and therefore do not require a linkage mechanism to convert reciprocating to rotary motion. Thus, they produce smoother rotational forces on the output shaft. This contributes to a lower maintenance requirement and less wear on the machinery they power than a comparable reciprocating engine.

The *Turbinia* - the first steam turbine-powered ship

The main use for steam turbines is in electricity generation (about 80% of the world's electric production is by use of steam turbines) and to a lesser extent as marine prime movers. In the former, the high speed of rotation is an advantage, and in both cases the relative bulk is not a disadvantage; in the latter (pioneered on the Turbinia), the light weight, high efficiency and high power are highly desirable.

Virtually all nuclear power plants and some nuclear submarines, generate electricity by heating water to provide steam that drives a turbine connected to an electrical generator for main propulsion. A limited number of steam turbine railroad locomotives were manufactured. Some non-condensing direct-drive locomotives did meet with some success for long haul freight operations in Sweden, but were not repeated. Elsewhere, notably in the U.S.A., more advanced designs with electric transmission were built experimentally, but not reproduced. It was found that steam turbines were not ideally suited to the railroad environment and these locomotives failed to oust the classic reciprocating steam unit in the way that modern diesel and electric traction has done.

Rotary steam engines

It is possible to use a mechanism based on a pistonless rotary engine such as the Wankel engine in place of the cylinders and valve gear of a conventional reciprocating steam engine. Many such engines have been designed, from the time of James Watt to the present day, but relatively few were actually built and even fewer went into quantity production; see link at bottom of article for more details. The major problem is the difficulty of sealing the rotors to make them steam-tight in the face of wear and thermal expansion; the resulting leakage made them very inefficient. Lack of expansive working, or any means of control of the cutoff is also a serious problem with many such designs. By the 1840s it was clear that the concept had inherent problems and rotary engines were treated with some derision in the technical press. However, the arrival of electricity on the scene, and the obvious advantages of driving a dynamo directly from a high-speed engine, led to something of a revival in interest in the 1880s and 1890s, and a few designs had some limited success.

Of the few designs that were manufactured in quantity, those of the Hult Brothers Rotary Steam Engine Company of Stockholm, Sweden, and the spherical engine of Beauchamp Tower are notable. Tower's engines were used by the Great Eastern Railway to drive lighting dynamos on their locomotives, and by the Admiralty for driving dynamos on board the ships of the Royal Navy. They were eventually replaced in these niche applications by steam turbines.

Jet type

Invented by Australian engineer Alan Burns and developed in Britain by engineers at Pursuit Dynamics, this underwater jet engine uses high pressure steam to draw in water through an intake at the front and expel it at high speed through the rear. When steam condenses in water, a shock wave is created and is focused by the chamber to blast water out of the back. To improve the engine's efficiency, the engine draws in air through a vent ahead of the steam jet, which creates air bubbles and changes the way the steam mixes with the water.

Unlike in conventional steam engines, there are no moving parts to wear out, and the exhaust water is only several degrees warmer in tests. The engine can also serve as pump and mixer. This type of system is referred to as 'PDX Technology' by Pursuit Dynamics.

Rocket type

The aeolipile represents the use of steam by the rocket-reaction principle, although not for direct propulsion.

In more modern times there has been limited use of steam for rocketry—particularly for rocket cars. The technique is simple in concept, simply fill a pressure vessel with hot water at high pressure, and open a valve leading to a suitable nozzle. The drop in pressure immediately boils some of the water and the steam leaves through a nozzle, giving a significant propulsive force.

It might be expected that water in the pressure vessel should be at high pressure; but in practice the pressure vessel has considerable mass, which reduces the acceleration of the vehicle. Therefore a much lower pressure is used, which permits a lighter pressure vessel, which in turn gives the highest final speed.

There are even speculative plans for interplanetary use. Although steam rockets are relatively inefficient in their use of propellant, this very well may not matter as the solar system is believed to have extremely large stores of water ice which can be used as propellant. Extracting this water and using it in interplanetary rockets requires several orders of magnitude less equipment than breaking it down to hydrogen and oxygen for conventional rocketry.

Advantages

The strength of the steam engine for modern purposes is in its ability to convert heat from almost any source into mechanical work. Unlike the internal combustion engine, the steam engine is not particular about the source of heat. Most notably, without the use of a steam engine it would be more difficult to harness nuclear energy for useful work, as a nuclear reactor does not directly generate either mechanical work or electrical energy—the reactor itself simply heats or boils water. It is the steam engine which converts the heat energy into useful work. Steam may also be produced without combustion of fuel, through solar concentrators. A demonstration power plant has been built using a central heat collecting tower and a large number of solar tracking mirrors, (called heliostats). (see Whitecliffs Project)

Similar advantages are found in a different type of external combustion engine, the Stirling engine, which can offer efficient power (with advanced regenerators and large radiators) at the cost of a much lower power-to-size/weight ratio than even modern steam engines with compact boilers. These Stirling engines are not commercially produced, although the concepts are promising.

Steam locomotives are especially advantageous at high elevations as they are not adversely affected by the lower atmospheric pressure. This was inadvertently discovered when steam locomotives operated at high altitudes in the mountains of South America were replaced by diesel-electric units of equivalent sea level power. These were quickly replaced by much more powerful locomotives capable of producing sufficient power at high altitude.

In Switzerland (Brienz Rothhorn) and Austria (Schafberg Bahn) new rack steam locomotives have proved very successful. They were designed based on a 1930s design of Swiss Locomotive and Machine Works (SLM) but with all of today's possible improvements like roller bearings, heat insulation, light-oil firing, improved inner streamlining, one-man-driving and so on. These resulted in 60 percent lower fuel consumption per passenger and massively reduced costs for maintenance and handling. Economics now are similar or better than with most advanced diesel or electric systems. Also a steam train with similar speed and capacity is 50 percent lighter than an electric or diesel train, thus, especially on rack railways, significantly reducing wear and tear on the track. Also, a new steam engine for a paddle steam ship on Lake Geneva, the *Montreux*, was designed and built, being the world's first full-size ship steam engine with an electronic remote control . The steam group of SLM in 2000 created a wholly-owned company called DLM to design modern steam engines and steam locomotives.

Safety

Steam engines possess boilers and other components that are pressure vessels that contain a great deal of potential energy. Steam explosions can and have caused great loss of life in the past. While variations in standards may exist in different countries, stringent legal, testing, training, care with manufacture, operation and certification is applied to try to minimise or prevent such occurrences.

Failure modes include:

- overpressurisation of the boiler
- insufficient water in the boiler causing overheating and vessel failure
- pressure vessel failure of the boiler due to inadequate construction or maintenance.
- escape of steam from pipework/boiler causing scalding

Efficiency

The efficiency of an engine can be calculated by dividing the number of joules of mechanical work that the engine produces by the number of joules of energy input to the engine by the burning fuel. The rest of the energy is dumped into the environment as heat.

No pure heat engine can be more efficient than the Carnot cycle, in which heat is moved from a high temperature reservoir to one at a low temperature, and the efficiency depends on the temperature difference. Hence, steam engines should ideally be operated at the highest steam temperature possible (superheated steam), and release the waste heat at the lowest temperature possible.

In practice, a steam engine exhausting the steam to atmosphere will have an efficiency (including the boiler) of 1% to 8%, but with the addition of a condenser and multiple expansion engines the efficiency may be greatly improved to 25% or better. A power station with steam reheat, etc. will achieve 30% to 42% efficiency. Combined cycle in which the burning material is first used to drive a gas turbine can produce 50% to 60% efficiency. It is also possible to capture the waste heat using cogeneration in which the residual steam is used for heating. It is therefore possible to use as much as 90% of the energy produced by burning fuel—only 10% of the energy produced by the combustion of the fuel goes wasted into the atmosphere.

The reason for varying efficiencies is because of the thermodynamic rule of the Carnot Cycle. The efficiency is the absolute temperature of the cold reservoir over the absolute temperature of the steam, subtracted from one. As the temperature changes in seasons, the efficiency changes with it, unless the cold reservoir is kept in an isothermal state. It should be noted that the Carnot Cycle calculations **require** absolute temperatures.

One source of inefficiency is that the condenser causes losses by being somewhat hotter than the outside world, although this can be mitigated by condensing the steam in a heat exchanger and using the recovered heat, for example to pre-heat the air being used in the burner of an external combustion engine.

The operation of the engine portion alone is not dependent upon steam; any pressurized gas may be used. Compressed air is sometimes used to test or demonstrate small model "steam" engines.

Television

2008/9 Schools Wikipedia Selection. Related subjects: Engineering; Television

Television (often abbreviated to **TV**) is a widely used telecommunication system for broadcasting and receiving moving pictures and sound over a distance. The term may also be used to refer specifically to a television set, programming or television transmission. The word is derived from mixed Latin and Greek roots, meaning "far sight": Greek *tele* (τῆλε), far, and Latin *vision*, sight (from *video, vis-* to see, or to view in the first person).

Braun HF 1, Germany, 1959

Since it first became commercially available from the late 1930s, the television set has become a common household communications device in homes and institutions, particularly in the First World, as a source of entertainment and news. Since the 1970s, video recordings on VCR tapes and later, digital playback systems such as DVDs, have enabled the television to be used to view recorded movies and other programs.

A television system may be made up of multiple components, so a screen which lacks an internal tuner to receive the broadcast signals is called a monitor rather than a television. A television may be built to receive different broadcast or video formats, such as high-definition television, commonly referred to as HDTV. HDTV costs more than normal TV but is becoming more available.

Clivia II FER858A (VEB Rafena, Radeberg, Germany), 1956

History

Technology

Geographical usage

- Timeline of the introduction of television in countries

Content

Programming

Getting TV programming shown to the public can happen in many different ways. After production the next step is to market and deliver the product to whatever markets are open to using it. This typically happens on two levels:

1. **Original Run** or **First Run** – a producer creates a program of one or multiple episodes and shows it on a station or network which has either paid for the production itself or to which a license has been granted by the producers to do the same.
2. **Syndication** – this is the terminology rather broadly used to describe secondary programming usages (beyond original run). It includes secondary runs in the country of first issue, but also international usage which may or may not be managed by the originating producer. In many cases other companies, TV stations or individuals are engaged to do the syndication work, in other words to sell the product into the markets they are allowed to sell into by contract from the copyright holders, in most cases the producers.

In most countries, the first wave occurs primarily on free-to-air (FTA) television, while the second wave happens on subscription TV and in other countries. In the U.S., however, the first wave occurs on the FTA networks and subscription services, and the second wave travels via all means of distribution.

First run programming is increasing on subscription services outside the U.S., but few domestically produced programs are syndicated on domestic FTA elsewhere. This practice is increasing however, generally on digital-only FTA channels, or with subscriber-only first run material appearing on FTA.

Unlike the U.S., repeat FTA screenings of a FTA network program almost only occur on that network. Also, Affiliates rarely buy or produce non-network programming that is not centred around local events.

Funding

Advertising

- United States

Since inception in the U.S. in 1940, TV commercials have become one of the most effective, persuasive, and popular method of selling products of many sorts, especially consumer goods. U.S. advertising rates are determined primarily by Nielsen Ratings. The time of the day and popularity of the channel determine how much a television commercial can cost. For example, the highly popular American Idol can cost approximately $750,000 for a thirty second block of commercial time; while the same amount of time for the World Cup and the Super Bowl can cost several million dollars.

In recent years, the paid program or infomercial has become common, usually in lengths of 30 minutes or one hour. Some drug companies have even created "news" items for broadcast, paying program directors to use them.

Some TV programs also weave advertisements into their shows, a practice begun in film and known as product placement. For example, a character could be drinking a certain kind of soda, going to a particular chain restaurant, or driving a certain make of car. (This is sometimes very subtle, where shows have vehicles provided by manufacturers for low cost, rather than wrangling them.) Sometimes a specific brand or trade mark, or music from a certain artist or group, is used. (This excludes guest appearances by artists, who perform on the show.)

- United Kingdom

The TV regulator oversees the TV advertising in the United Kingdom. Its restrictions have applied since the early days of commercially funded TV in the UK. Despite this, the demand from advertisers ensured that ownership of a commercial broadcasting licence was, at one time, likened by the TV mogul, Lew Grade, as a being a "licence to print money". The restrictions mean that the big three national commercial TV channels, ITV, Channel 4, and Five can show an average of only seven minutes of advertising per hour (eight minutes in the peak period). Other broadcasters must average no more than nine minutes (twelve in the peak). This means that many imported TV shows from the US have un-natural breaks where the UK company has edited out the breaks intended for US advertising. Advertisements must not be inserted in the course of any broadcast of a news or current affairs program of less than half an hour scheduled duration, or in a documentary of less than half an hour scheduled duration, or in a program for children of less than half an hour scheduled duration. Nor may advertisements be carried in a program designed and broadcast for reception in schools or in any religious service or other devotional program, or during a formal Royal ceremony or occasion. There also must be clear demarcations in time between the programs and the advertisements. The BBC, being strictly non-commercial is not allowed to show advertisements on television, the majority of its budget comes from TV licencing (see below).

Taxation or **TV License**

Television services in some countries may be funded by a television licence, a form of taxation which means advertising plays a lesser role or no role at all. For example, in the United Kingdom, and in many European countries, some channels may carry no advertising at all and some very little. The British Broadcasting Corporation (BBC) carries no advertising and is funded by an annual licence paid by all households owning a television. This licence fee is set by government, but the BBC is not answerable to or controlled by government and is therefore genuinely independent. The fee also funds radio channels, transmitters and the BBC.co.uk online service. Advertising has been introduced to the internationally-facing BBC.com website in a small way to fund broadband content delivered outside of the United Kingdom.

Subscription

Some TV channels are partly funded from subscriptions and therefore the signals are encrypted before broadcast to ensure that only paying subscribers have access to the decryption codes. Some subscription services are also funded by advertising.

Television genres

Television genres include a broad range of programming types that entertain, inform, and educate viewers. The most expensive entertainment genres to produce are usually drama and dramatic miniseries. However, other genres, such as historical Western genres, may also have high production costs.

Popular entertainment genres include action-oriented shows such as police, crime, detective dramas, horror or thriller shows. As well, there are also other variants of the drama genre, such as medical dramas and daytime soap operas. Sci-fi (Science fiction) shows can fall into either the drama category or the action category, depending on whether they emphasize philosophical questions or high adventure. Comedy is a popular genre which includes sitcoms (Situation Comedy) and animated shows for the adult demographic such as *Family Guy*.

The least expensive forms of entertainment programming are game shows, talk shows, variety shows, and reality TV. Game shows show contestants answering questions and solving puzzles to win prizes. Talk shows feature interviews with film, television and music celebrities and public figures. Variety shows feature a range of musical performers and other entertainers such as comedians and magicians introduced by a host or Master of Ceremonies. There is some crossover between some talk shows and variety shows, because leading talk shows often feature performances by bands, singers, comedians, and other performers in between the interview segments.

Reality TV shows show "regular" people (i.e., not actors) who are facing unusual challenges or experiences, ranging from arrest by police officers (*COPS*) to weight loss (*The Biggest Loser*). A variant version of reality shows depicts celebrities doing mundane activities such as going about their everyday life (*The Osbournes*) or doing manual labour jobs (*Simple Life*).

One of the television genres, the children's and youth genre is defined by the audience, rather than by the content of the programming. Children's programming includes animated programs aimed at the child demographic, documentaries for children, and music/variety shows targeted at kids. There is overlap between the children's/youth genre and other genres, such as the educational genre.

Social aspects

Television has played a pivotal role in the socialization of the 20th and 21st centuries. There are many social aspects of television that can be addressed, including:

- Positive effects
- Negative effects
- Gender and television
- Politics and television
- Socializing children
- Technology trends
- Suitability for audience
- Alleged dangers
- Propaganda delivery
- Educational advantages

Environmental aspects

With high lead content in CRTs, and the rapid diffusion of new, flat-panel display technologies, some of which (LCDs) use lamps containing mercury, there is growing concern about electronic waste from discarded televisions. Related occupational health concerns exist, as well, for disassemblers removing copper wiring and other materials from CRTs. Further environmental concerns related to television design and use relate to the devices' increasing electrical energy requirements. Some speculate that television is responsible for a "dumbing down" of modern peoples. Several articles have been written by numerous individuals, and watchgroups. One theory is that the passive state of the brain leads to a form of mental atrophy. While watching television the higher functions of the brain slow down and in some cases cease. There are ongoing studies to determine the risk of this passive activity.

Train

2008/9 Schools Wikipedia Selection. Related subjects: Railway transport

Rail transport
Operations
Stations
Trains
Locomotives
Rolling stock
History
History by country
Terminology
By country
Accidents

Modelling

A **train** is a connected series of vehicles that move along a track (permanent way) to transport freight or passengers from one place to another. The track usually consists of two rails, but might also be a monorail or maglev guideway. Propulsion for the train is provided by a separate locomotive, or from individual motors in self-propelled multiple units. Most modern trains are powered by diesel locomotives or by electricity supplied by overhead wires or additional rails, although historically (from the early 19th century to the mid-20th century) the steam locomotive was the dominant form of locomotive power. Other sources of power (such as horses, rope or wire, gravity, pneumatics, and gas turbines) are possible.

The word 'train' comes from the Old French *trahiner*, itself from the Latin *trahere* 'pull, draw'.

Types of trains

There are various types of train designed for particular purposes. A train can consist of a combination of one or more locomotives and attached railroad cars, or a self-propelled multiple unit (or occasionally a single powered coach, called a railcar). Trains can also be hauled by horses, pulled by a cable, or run downhill by gravity.

An SP freight train west of Chicago in 1992.

Special kinds of trains running on corresponding special 'railways' are atmospheric railways, monorails, high-speed railways, maglev, rubber-tired underground, funicular and cog railways.

A passenger train may consist of one or several locomotives, and one or more coaches. Alternatively, a train may consist entirely of passenger carrying coaches, some or all of which are powered as a " multiple unit". In many parts of the world, particularly Japan and Europe, high-speed rail is utilized extensively for passenger travel.

German ICE high speed train

Freight trains comprise wagons or trucks rather than carriages, though some parcel and mail trains (especially Travelling Post Offices) are outwardly more like passenger trains.

Trains can also be 'mixed', comprising both passenger accommodation and freight vehicles. Such mixed trains are most likely to occur where services are infrequent, and running separate passenger and freight trains is not cost-effective, though the differing needs of passengers and freight usually means this is avoided where possible.

Special trains are also used for track maintenance; in some places, this is called maintenance of way.

In the United Kingdom, a train hauled by two locomotives is said to be "double-headed", and in Canada and the United States it is quite common for a long freight train to be headed by three or more locomotives. A train with a locomotive attached at each end is described as 'top and tailed', this practice typically being used when there are no reversing facilities available. Where a second locomotive is attached temporarily to assist a train up steep banks or grades (or down them by providing braking power) it is referred to as ' banking' in the UK, or 'helper service' in North America. Recently, many loaded trains in the US have been made up with one or more locomotives in the middle or at the rear of the train, operated remotely from the lead cab. This is referred to as "DP" or "Distributed Power."

Official terminology

The railway terminology that is used to describe a 'train' varies between countries.


A British Rail Class 153 DMU

United Kingdom

In the United Kingdom, the interchangeable terms **set** and **unit** are used to refer to a group of permanently or semi-permanently coupled vehicles, such as those of a multiple unit. While when referring to a train made up of a variety of vehicles, or of several sets/units, the term **formation** is used. (Although the UK public and media often forgo 'formation', for simply 'train'.) The word **rake** is also used for a group of coaches or wagons.

In the United Kingdom Section 83(1) of the Railways Act 1993 defines "train" as follows:

> a) two or more items of rolling stock coupled together, at least one of which is a locomotive; or
> b) a locomotive not coupled to any other rolling stock.


Modern German Class 423 EMU trainsets meet each other

United States

In the United States, the term **consist** is used to describe the group of rail vehicles which make up a train. When referring to motive power, **consist** refers to the group of locomotives powering the train. Similarly, the term **train set** refers to a group of rolling stock that is permanently or semi-permanently coupled together to form a unified set of equipment (the term is most often applied to passenger train configurations). Also, in the United States, they sometimes call the engine an 'iron horse', but varies by person as well. The term 'iron horse' was thought of when the steam locomotive first appeared in the United States. They called it that, due to the fact that it replaced the horse on the railway lines, and was made of metal.

The Atchison, Topeka and Santa Fe Railway's 1948 operating rules define a train as: "An engine or more than one engine coupled, with or without cars, displaying markers."

Motive power

The first trains were rope-hauled, gravity powered or pulled by horses, but from the early 19th century almost all were powered by steam locomotives. From the 1920s onwards they began to be replaced by less labour intensive and cleaner (but more complex and expensive) diesel locomotives and electric locomotives, while at about the same time self-propelled multiple unit vehicles of either power system became much more common in passenger service. In most countries dieselisation of locomotives in day-to-day use was completed by the 1970s. A few countries, most notably the People's Republic of China, where coal and labour are cheap, still use steam locomotives, but this is being gradually phased out. Historic steam trains still run in many other countries, for the leisure and enthusiast market.

V43, a common Hungarian electric locomotive used in passenger train service.

Electric traction offers a lower cost per mile of train operation but at a higher initial cost, which can only be justified on high traffic lines. Since the cost per mile of construction is much higher, electric traction is less favored on long-distance lines with the exception of long-distance high speed lines. Electric trains receive their current via overhead lines or through a third rail electric system.

Passenger trains

A passenger train is one which includes passenger-carrying vehicles. It may be a self-powered multiple unit or railcar, or else a combination of one or more locomotives and one or more unpowered trailers known as coaches, cars or carriages. Passenger trains travel between stations where passengers may join or leave the train. Many of the more prestigious passenger train services have been given a specific name, some of which have become famous in literature and fiction. India has the largest passenger density in the world.India has one of the largest passenger density due to a great population, refering to a population chart of India, yet it is only one of the most populated countries, but is the only populated one to have the greatest passenger density out of the other countries.

Long-distance trains

Long-distance trains travel between many cities and/or regions of a country, and sometimes cross several countries. They often have a dining car or restaurant car to allow passengers to have a meal during the course of their journey. Trains traveling overnight may also have sleeping cars.

Interior of a passenger car in a long-distance train in Finland

High-speed trains

High speed trains normally travel during the day. They compete with airliners in speed. In Japan, most of the public transportation between the Tokyo metropolitan area and the Osaka metropolitan area (with around 500 km in distance between them) is dominated by the Shinkansen, however in travel further than around 500 km (such as Tokyo- Hiroshima) more people prefer to travel by air.

Japanese Shinkansen 500 Series (High-speed rail)

Very fast trains sometimes tilt, like the APT, the Pendolino, or the Talgo. Tilting is a system where the passenger cars automatically lean into curves, reducing the sideways g-forces on passengers and permitting higher speeds on curves in the track with greater passenger comfort.

The fastest train on rails is the French TGV (Train à Grande Vitesse) (French for High Speed Train) which achieved a 574.8 km/h (356 mph) speed in testing in 2007. However, TGVs run at a maximum commercial speed of 300-320 km/h. The German ICE uses this commercial speed of 300-320 km/h as well.

Inter-city trains

Trains connecting cities can be distinguished into two groups, inter-city trains, which do not halt at small stations, and trains that serve all stations, usually known as local trains or "stoppers" (and sometimes an intermediate type, usually known as limited-stop).

Branch line trains

Branch lines are usually defined as connections to local stations or local lines and usually stopping services, running to all stations or the majority of stations on a line.

Commuter trains

For shorter distances many cities have networks of commuter trains, serving the city and its suburbs. Some carriages may be laid out to have more standing room than seats, or to facilitate the carrying of prams, cycles or wheelchairs. Some countries have double-decked passenger trains for use in conurbations. Double deck high speed and sleeper trains are becoming more common in mainland Europe.

The Mumbai Suburban Railway in India has the largest passenger density in the world

Passenger trains usually have emergency brake handles (or a "communication cord") that the public can operate. Abuse is punished by a heavy fine.

Large cities often have a metro system, also called underground, subway or tube. The trains are electrically powered, usually by third rail, and their railroads are separate from other traffic, without level crossings. Usually they run in tunnels in the city centre and sometimes on elevated structures in the outer parts of the city. They can accelerate and decelerate faster than heavier, long-distance trains.

Interior of a 6 door passenger car in Japan, when the bench seats are folded

A light one- or two-car rail vehicle running through the streets is by convention not considered a train but rather a tram, trolley, light-rail vehicle or streetcar, but the distinction is not always strict. In some countries such as the United Kingdom the distinction between a tramway and a railway is precise and defined in law.

The term light rail is sometimes used for a modern tram, but it may also mean an intermediate form between a tram and a train, similar to metro except that it may have level crossings. These are often protected with crossing gates. They may also be called a trolley.

Maglev trains and monorails represent minor technologies in the train field.

The term **rapid transit** is used for public transport such as commuter trains, metro and light rail. However, in New York City, lines on the New York City Subway have been referred to as "trains".

Some commuter trains in Tokyo, Japan have special cars which the bench seats fold up to provide standing room only during the morning rush hour (until 10 a.m.). The E231 series train has two of these cars in each set (usually as part of a 10- or 11-car set), officially nicknamed "roku-tobira-sha" (literally, "6 door car") - all the other cars have four sets of doors on each side.

An estimated 3.5 million passengers ride every day on Tokyo's Yamanote Line, with its 29 stations. For comparison, the New York City Subway carries 4.8 million passengers per day on 26 lines serving 468 stations.

Named trains

Railway companies often give a name to a train service as a marketing exercise, to raise the profile of the service and hence attract more passengers (and also to gain kudos for the company). Usually, naming is reserved for the most prestigious trains: the high-speed express trains between major cities, stopping at few intermediate stations. The names of services such as the Orient Express, the Flying Scotsman, the Flèche d'Or and the Royal Scot have passed into popular culture.

A somewhat less common practice is the naming of freight trains, for the same commercial reasons. The "Condor" was an overnight London-Glasgow express goods train, in the 1960s, hauled by pairs of "Metrovick" diesel locomotives. In the mid-1960s, British Rail introduced the "Freightliner" brand, for the new train services carrying containers between dedicated terminals around the rail network. The Rev. W. Awdry also named freight trains, coining the term *The Flying Kipper* for the overnight express fish train that appeared in his stories in The Railway Series books.

Freight trains

British electric container freight train

A freight train (also known as goods train) uses **freight cars** (also known as wagons or trucks) to transport goods or materials (cargo) – essentially any train that is not used for carrying passengers. Much of the world's freight is transported by train, and in the USA the rail system is used more for transporting freight than passengers.

Under the right circumstances, transporting freight by train is highly economic, and also more energy efficient than transporting freight by road. Rail freight is most economic when freight is being carried in bulk and over long distances, but is less suited to short distances and small loads. Bulk aggregate movements of a mere twenty miles (32 km) can be cost effective even allowing for trans-shipment costs. These trans-shipment costs dominate in many cases and many modern practices such as container freight are aimed at minimizing these.

American freight service

The main disadvantage of rail freight is its lack of flexibility. For this reason, rail has lost much of the freight business to road competition. Many governments are now trying to encourage more freight onto trains, because of the benefits that it would bring.

There are many different types of freight trains, which are used to carry many different kinds of freight, with many different types of wagons. One of the most common types on modern railways are container trains, where containers can be lifted on and off the train by cranes and loaded off or onto trucks or ships.

This type of freight train has largely superseded the traditional boxcar (wagon-load) type of freight train, with which the cargo has to be loaded or unloaded manually.

In some countries " piggy-back" trains are used: trucks can drive straight onto the train and drive off again when the end destination is reached. A system like this is used through the Channel Tunnel between England and France, and for the trans-Alpine service between France and Italy (this service uses Modalohr road trailer carriers). 'Piggy-back' trains are the fastest growing type of freight trains in the United States, where they are also known as ' trailer on flatcar' or TOFC trains. 'Piggy-back' trains require no special modifications to the vehicles being carried. An alternative type of "inter-modal" vehicle, known as a Roadrailer, is designed to be physically attached to the train. The original trailers were fitted with two sets of wheels: one set flanged, for the trailer to run connected to other such trailers as a rail vehicle in a train; and one set tyred, for use as the semi-trailer of a road vehicle. More modern trailers have only road wheels and are designed to be carried on specially adapted bogies (trucks) when moving on rails.

There are also many other types of wagons, such as "low loader" wagons for transporting road vehicles. There are refrigerator cars for transporting foods such as ice cream. There are simple types of open-topped wagons for transporting minerals and bulk material such as coal, and tankers for transporting liquids and gases. Today however most coal and aggregates are moved in hopper wagons that can be filled and discharged rapidly, to enable efficient handling of the materials.

Freight trains are sometimes illegally boarded by passengers who do not wish to pay money, or do not have the money to travel by ordinary means. This is referred to as " hopping" and is considered by some communities to be a viable form of transport. Most hoppers sneak into train yards and stow away in boxcars. More bold hoppers will catch a train "on the fly", that is, as it is moving, leading to occasional fatalities.

Tram

2008/9 Schools Wikipedia Selection. Related subjects: Road transport

A **tram**, **tramcar**, **trolley**, **trolley car**, or **streetcar** is a railborne vehicle, lighter than a train, designed for the transport of passengers (and/or, very occasionally, freight) within, close to, or between villages, towns and/or cities, primarily on streets.

Tram systems (or "tramways" or "street railways") were common throughout the industrialized world in the late 19th and early 20th centuries, but they disappeared from many cities in the mid-20th century. In recent years, they have made a comeback. Many newer light rail systems share features with trams, although a distinction is usually drawn between the two, especially if the line has significant off-street running.

Rail transport
Operations
Stations
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History
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Modelling

Use of the term

The terms "tram" and "tramway" were originally Scots and Northern English words for the type of truck used in coal mines and the tracks on which they ran — probably derived from the North Sea Germanic word "trame" of unknown origin meaning the "beam or shaft of a barrow or sledge", also "a barrow" or container body.

UK Metrolink Tram in Manchester, England

Although "tram" and "tramway" have been adopted by many languages, they are not used universally in English, North Americans preferring "trolley", "trolley car" or "streetcar". The term "streetcar" is first recorded in 1860, and is a North American usage, as is "trolley," which is believed to derive from the "troller," a four wheeled device that was dragged along dual overhead wires by a cable that connected the troller to the top of the car and collected electrical power from the overhead wire, sometimes simply strung, sometimes on a catenary. The trolley pole, which supplanted the troller early-on, is fitted to the top of the car and is spring-loaded in order to keep the trolley wheel, at the upper of the pole, firmly in contact with the overhead wire. The terms trolley pole and trolley wheel both derive from the troller.

Image:Zg tram1.jpg

Croatian TMK 2200 in Zagreb, made by: CROTRAM, low-floor tram

Modern trolleys often do not use a trolley wheel: either they have a metal shoe with a carbon insert or they dispense with the trolley pole completely and have instead a pantograph. Other streetcars are sometimes called trolleys, even though strictly this may be incorrect: cable cars, for example, or conduit cars that draw power from an underground supply.

Tourist buses made to look like streetcars are also sometimes called trolleys; see tourist trolley. Likewise, open, low-speed segmented vehicles on rubber tires, generally used to ferry tourists short distances, can be called trams, particularly in the U.S.; a famous example is the tram on the Universal Studios tour.

Electric buses, which still overwhelmingly use twin trolley poles (one for live current, one for return) are called **trolleybuses**, **trackless trolleys** (particularly in the U.S.), or sometimes also **trolleys**.

History

TW2000 car in Hanover, Germany

Volkswagen Cargo-Tram in Dresden, Germany on a section of grassed track. It delivers parts to the Transparent Factory

Image:Frankfurttramgermany.

Public transport in Frankfurt am Main Trams in Frankfurt, Germany

The very first tram (streetcar) was the Mumbles Railway (Swansea to Mumbles, Wales, UK) it was horse drawn at first and later by steam power and then electric. The Mumbles Railway Act 1804 was passed by the British Parliament, and the first passenger railway (which acted like streetcars did in the US some 30 years later) started operating in 1807.

The first streetcars, also known as horsecars in North America, were built in the United States and developed from city stagecoach lines and omnibus lines that picked up and dropped off passengers on a regular route and without the need to be pre-hired. These trams were an animal railway, usually using horses and sometimes mules to haul the cars, usually two as a team. Rarely other animals were tried, including humans in emergencies. The first streetcar - the New York and Harlem Railroad's Fourth Avenue Line - ran along the Bowery and Fourth Avenue in New York City, and began service in the year 1832. It was followed in 1835 by New Orleans, Louisiana, which is the oldest continuously operating street railway system in the world, according to the American Society of Mechanical Engineers. At first the rails protruded above street level, causing accidents and major trouble for pedestrians. They were supplanted in 1852 by grooved rails or girder rails, invented by Alphonse Loubat. The first tram in Paris, France, was inaugurated in 1853 for the upcoming World's Fair, where a test line was presented along the Cours de la Reine, in the 8th arrondissement.

One of the advantages over earlier forms of transit was the low rolling resistance of metal wheels on steel rails, allowing the animals to haul a greater load for a given effort. Problems included the fact that any given animal could only work so many hours on a given day, had to be housed, groomed, fed and cared for day in and day out, and produced prodigious amounts of manure, which the streetcar company was charged with disposing of. Since a typical horse pulled a car for perhaps a dozen miles a day and worked for four or five hours, many systems needed ten or more horses in stable for each horsecar. Electric trams largely replaced animal power in the late 19th and early 20th century. New York City had closed its last horsecar line in 1917. The last regular mule drawn streetcar in the U.S.A., in Sulphur Rock, Arkansas, closed in 1926. However during World War II some old horse cars were temporarily returned to service to help conserve fuel. A mule-powered line in Celaya, Mexico, operated until 1956. Horse-drawn trams still operate in Douglas, Isle of Man. There is also a small line operated on Main Street at DisneyWorld, outside of Orlando Florida. A small horse-drawn service operates every 40 minutes at Victor Harbour, South Australia, daily with 20 minute services during tourist seasons. This service runs between the mainland and Granite Island across a causeway.

The tram developed after that in numerous cities of Europe (London, Berlin, Paris, etc.) and Asia (Kyoto, Tokyo, Hong Kong). Faster and more comfortable than the omnibus, trams had a high cost of operation because they were pulled by horses. That is why mechanical drives were rapidly developed, with steam power in 1873, and electrical after 1881, when Siemens AG presented the electric drive at the International Electricity Exhibition in Paris.

The convenience and economy of electricity resulted in its rapid adoption once the technical problems of production and transmission of electricity were solved. The first prototype of the electric tram was developed by Russian engineer Fyodor Pirotsky. He modified a Horse tramway car to be powered by electricity instead of horses. The invention was tested in 1880 in Saint Petersburg, Russia. The world's first electric tram line opened in Lichterfelde near Berlin, Germany, in 1881. It was built by Werner von Siemens. (see Berlin Straßenbahn).

In Japan, the Kyoto Electric railroad was the first tram system, starting operation in 1865. By 1932, the network had grown to 82 railway companies in 65 cities, with a total network length of 1,479km. By the 1960s, however, the tram had generally died out in Japan.

History of the different types of tram

Horse-drawn trams

In the nineteenth century Calcutta (now Kolkata) was developing fast as a British trading and business centre. Transport was mainly by palanquins carried on men's shoulders, phaetons pulled by horses, etc. In 1867, The Calcutta Corporation, with financial assistance from the Government of Bengal developed mass transport. The first tramcar rolled out on the streets of Calcutta on February 24, 1873, with horse drawn coaches running on steel rails between Sealdah and Armenian Ghat via Bowbazar and Dalhousie Square, (now B. B. D. Bagh). The Corporation entered into an agreement on February 10, 1879 with three English industrial magnates: Robinson Soutter, Alfred Parrish and Dilwyn Parrish. Registered in London, the Calcutta Tramways Company came into existence in 1880 after the sanction of The Calcutta Tramways Act, 1880.

Horse drawn trams in Calcutta (now Kolkata), India - Life size model at City Centre arcade

By 1902 Messrs Kilburn & Co completed the electrification of the Calcutta tramways and the first electric tramcar was introduced in the Kidderpore section.

Calcutta remains the only Indian city which has maintained tramway system. As of now, it remains an unreliable but very comfortable and eco-friendly transport.

A horse tramway in Gdańsk, Poland (late 19th century)

Steam trams

The first mechanical trams were operated using mobile steam engines. Generally, there were two types of steam tram. The first and most common had a small steam locomotive (called a tram engine in the UK) at the head of a line of one or more carriages, similar to a small train. Systems with such steam trams included Christchurch, New Zealand, Sydney, Australia, and other provincial city systems in New South Wales.


Steam trams in Rockhampton, Queensland - note the small boiler at the front of the leading tram.

The other style of steam tram had the steam engine mounted in the body of the tram. The most notable system to adopt such trams was in Paris. French-designed steam trams also operated in Rockhampton, in the Australian state of Queensland between 1909 and 1939. Stockholm, Sweden, also had a steam tramline at the island of Södermalm between 1887 and 1901. A major drawback of this style of tram was the limited space for the engine, so that these trams were usually underpowered.

Cable pulled cars

The next type of tram was the cable car, which sought to reduce labor costs and the hardship on animals. Cable cars are pulled along a rail track by a continuously moving cable running at a constant speed on which individual cars stop and start by releasing and gripping this cable as required. The power to move the cable is provided at a site away from the actual operation. The first cable car line in the United States was tested in San Francisco, California, in 1873. The second city to operate cable trams was Dunedin in New Zealand in 1881. Dunedin's cable trams ceased operation in 1957.

Cable cars suffered from high infrastructure costs, since a vast and expensive system of cables, pulleys, stationary engines and vault structures between the rails had to be provided. They also require strength and skill to operate, to avoid obstructions and other cable cars. The cable had to be dropped at particular locations and the cars coast, for example when crossing another cable line. Breaks and frays in the cable, which occurred frequently, required the complete cessation of services over a cable route, while the cable was repaired. After the development of electrically-powered trams, the more costly cable car systems declined rapidly.

Cable cars were especially useful in hilly cities, partially explaining their survival in San Francisco, though the most extensive cable system in the U.S. was in Chicago, a much flatter city. The largest cable system in the world which operated in the flat city of Melbourne, Victoria, Australia, had, at its peak, 592 trams running on 74 kilometres of track.

The San Francisco cable cars, though significantly reduced in number, continue to perform a regular transportation function, in addition to being a tourist attraction. Single lines also survive on hilly parts of Wellington, New Zealand (rebuilt in 1979 to a funicular system but still called the ' Wellington Cable Car') and Hong Kong.

Electric trams (trolley cars)

Multiple functioning experimental electric trams were exhibited at the 1884 World Cotton Centennial World's Fair in New Orleans, Louisiana; however they were deemed as not yet adequately perfected to replace the Lamm fireless engines then propelling the St. Charles Avenue Streetcar in that city.

1923 postcard of Richmond Virginia showing the mass produced Pearly Thomas electric-powered trolley streetcars commonly seen in most US cities in the era.

Electric-powered trams (trolley cars, so called for the trolley pole used to gather power from an unshielded overhead wire), were first successfully tested in service in Richmond, Virginia, in 1888, in the Richmond Union Passenger Railway built by Frank J. Sprague. There were earlier commercial installations of electric streetcars, including one in Berlin, as early as 1881 by Werner von Siemens and the company that still bears his name, and also one in Saint Petersburg, Russia, invented and tested by Fyodor Pirotsky in 1880. Another was by John Joseph Wright, brother of the famous mining entrepreneur Whitaker Wright, in Toronto in 1883. The earlier installations, however, proved difficult and/or unreliable. Siemens' line, for example, provided power through a live rail and a return rail, like a model train setup, limiting the voltage that could be used, and providing unwanted excitement to people and animals crossing the tracks. Siemens later designed his own method of current collection, this time from an overhead wire, called the bow collector. Once this had been developed his cars became equal to, if not better than, any of Sprague's cars. The first electric interurban line connecting St. Catharines and Thorold, Ontario was operated in 1887, and was considered quite successful at the time. While this line proved quite versatile as one of the earliest fully functional electric streetcar installations, it still required horse-drawn support while climbing the Niagara Escarpment and for two months of the winter when hydroelectricity was not available. This line continued service in its original form well into the 1950s.

Tramways on ice of the River Neva in Saint Petersburg

Since Sprague's installation was the first to prove successful in all conditions, he is credited with being the inventor of the trolley car. He later developed Multiple unit control, first demonstrated in Chicago in 1897, allowing multiple cars to be coupled together and operated by a single motorman. This gave birth to the modern subway train.

A 1925 vintage British tram, a common sight until the 1950s

Two rare but significant alternatives were conduit current collection, which was widely used in London, Washington, D.C. and New York, and the Surface Contact Collection method, used in Wolverhampton (The Lorain System) and Hastings (The Dolter Stud System), UK.

Attempts to use on-board batteries as a source of electrical power were made from the 1880s and 1890s, with unsuccessful trials conducted (among other places) in Bendigo and Adelaide in Australia, although run for about 14 years as Hague *accutram* of HTM in the Netherlands.

A very famous Welsh example of a tram system was usually known as the Mumbles Train, or more formally as the Swansea and Mumbles Railway. Originally built as the Oystermouth Railway in 1804, on March 25, 1807 it became the first passenger-carrying railway in the world. Converted to an overhead cable-supplied system it operated electric cars from March 2, 1929 until its closure on January 5, 1960. These were the largest tram cars built for use in Britain and could each seat 106 passengers.

Another early tram system operated from 1886 until 1930 in Appleton, Wisconsin, and is notable for being powered by the world's first hydroelectric power station, which began operating on September 30, 1882 as the Appleton Edison Electric Company.

There is one particular hazard associated with trams powered from a trolley off an overhead line. Since the tram relies on contact with the rails for the current return path, a problem arises if the tram is derailed or (more usually) if it halts on a section of track that has been particularly heavily sanded by a previous tram, and the tram loses electrical contact with the rails. In this event, the main chassis of the tram, by virtue of a circuit path through ancillary loads (such as saloon lighting), is live by the full supply voltage (typically 600 volts) relative to the running rails (and indeed the surrounding earthed land). In British terminology such a tram was said to be 'grounded' - not to be confused with the US English use of the term which means the exact opposite. Any person stepping off the tram completed the earth return circuit and could receive a nasty electric shock. In such an event the driver was required to jump off the tram (avoiding simultaneous contact with the tram and the ground) and pull down the trolley before allowing passengers off the tram. Unless derailed, the tram could usually be recovered by running water down the running rails from a point higher than the tram. The water providing a conducting bridge between the tram and the rails.

Low floor

and Ultra Low Floor

A Bulgarian built T8M-900 tram with low floor middle section in Sofia.

The latest generation of LRVs has the advantage of partial or fully low-floor design, with the floor of the vehicles only 300 to 360 mm (12-14 inches) above top of rail, a capability not found in either rapid rail transit vehicles or streetcars. This allows them to load passengers, including ones in wheelchairs, directly from low-rise platforms that are not much more than raised sidewalks. This satisfies requirements to provide access to disabled passengers without using expensive wheelchair lifts, while at the same time making boarding faster and easier for other passengers as well. The City Class LRV (Citytram) is one example of a low floored vehicle, 300 mm above rail height, with 70% of the 29 m long and 75% of the 38 m long versions low floor. The low floor extends across the articulation. The City Class has been designed to operate around 15 m curves and climb 10% gradients, and therefore allow new systems to be built in existing urban streets without the need to demolish buildings.

Articulated

Articulated trams are tram cars that consist of several sections held together by flexible joints and a round platform. Like articulated buses, they have an increased passenger capacity. These trams can be up to forty metres in length, while a regular tram has to be much shorter. With this type, a Jacobs bogie supports the articulation between the two or more carbody sections. An articulated tram may be low floor variety or high (regular) floor variety. Since 1981 onwards, nearly 150 articulated LRV-trams of the last kind are e.g. to be found in The Hague Netherlands.

Ref.: HTM LRV nl:GTL8 / D.A. Borgdorff / The Hague - 2000 / ISBN 9090139354

Tram-train

Tram-train operation uses vehicles such as the Flexity Link and Regio- Citadis which are suited for use on urban tram lines, but also meet the necessary indication, power, and resistance requirements to be certified for operation on main line railways. This allows passengers to travel from suburban areas into city-centre destinations without having to change from a train to a tram when they arrive at the central station.

It has been primarily developed in Germanic countries, in particular Germany and Switzerland. Karlsruhe is a notable pioneer of the tram-train.

Cargo trams

Goods have been carried on rail vehicles through the streets, particularly near docks and steelworks, since the 19th century (most evident in Weymouth), and some Belgian *vicinale* routes were used to haul timber. At the turn of the 21st century, a new interest has arisen in using urban tramway systems to transport goods. The motivation now is to reduce air pollution, traffic congestion and damage to road surfaces in city centres. Dresden has a regular *CarGoTram* service, run by the world's longest tram trainsets (59.4 m), carrying car parts across the city centre to its Volkswagen factory. Vienna and Zürich use trams as mobile recycling depots. Kislovodsk had a freight-only tram system comprising one line which was used exclusively to deliver bottled Narzan mineral water to the railway station.

As of 7 March 2007, Amsterdam is piloting a cargo tram operation, which could reduce particulate pollution by 20% by halving the number of lorries – currently 5,000 - unloading in the inner city during the permitted timeframe from 07:00 till 10:30.

The pilot, operated by City Cargo Amsterdam, involves two cargo trams, operating from a distribution centre at Lutkemeerpolder, on the A9 ring motorway near the Osdorp terminus of tram no. 1. Each cargo tram can transport the load of 4 lorries (roughly 100 tonnes) to a ‘hub’ at Frederiksplein, where electric trucks deliver to the final destination.

If the trial is successful an investment of 100 million euro would see a fleet of 52 cargo trams distributing from four peripheral ‘cross docks’ to 15 inner-city hubs by 2012. These specially-built vehicles would be 30 metres long with 12 axles and a payload of 30 tonnes.

(Source: *Samenwest* 5 December 6, NOS3 television news 7 March 7)

Model trams

Models of trams are popular in HO scale and sometimes in 1:50 scale. They typically are powered and will accept plastic figures inside. Common manufacturers are Roco and Lima with many custom models being made as well. The German firm Hödl and the Austrian Halling specialize in trams in 1:87 scale.

A number of 1:76.2 scale tram models, especially kits, are made in the UK. Many of these run on 16.5 mm gauge track, which is incorrect for the representation of standard (4ft 8½ins) gauge, as it represents 4ft 1½ins in 4 mm (1:76.2) scale. This scale/gauge hybrid is called OO scale.

There are some Russian tram models available in 1:48 scale

Pros and cons of tram systems

All transit services involves a tradeoff between speed and frequency of stops. Services that stop frequently have a lower overall speed, and are therefore less attractive for longer trips. Metros, light rail, monorail, and bus rapid transit are all forms of rapid transit — which generally signifies high speed and widely-spaced stops. Trams are a form of local transit, making frequent stops. Thus, the most meaningful comparison of advantages and disadvantages is with other forms of local transit, primarily the local bus.

tram in Strasbourg, 2004.

Advantages

- Multiple entrances allow trams to load faster than suburban coaches, which tend to have a single entrance. This, combined with swifter acceleration and braking, lets trams maintain higher overall speeds than buses, if congestion allows.
- Trams can adapt to the number of passengers by adding additional cars during rush hour (as well as removing excess cars during off-peak hours). No additional driver is then required for the trip in comparison to buses.
- In general, trams provide a higher capacity service than buses.
- Unlike buses, but like trolleybuses, (electric) trams give off no exhaust emissions at point of use. Compared to motorbuses the noise of trams is generally perceived to be less disturbing.
- Rights-of-way for trams are narrower than for buses. This saves valuable space in cities with high population densities and/or narrow streets.
- Because they are rail-bound, trams command more respect from other road users than buses do, when operating on-road. In heavy traffic conditions, rogue drivers are less likely to hold up trams, for example by blocking intersections or parking on the road. This often leads to fewer delays. As a rule, especially in European cities and Melbourne, trams **always** have priority.
- Passenger comfort is normally superior to buses because of controlled acceleration and braking and curve easement. Rail transport such as used by trams provides a smoother ride than road use by buses.
- In most countries, trams don't suffer from the image problem that plagues buses. On the contrary — most people associate trams with a positive image. Unlike buses, trams tend to be popular with a wider spectrum of the public, including better-off people who often shun buses. This high level of customer acceptance means higher patronage and greater public support for investment in new tram infrastructure.

Disadvantages

- The capital cost is higher than for buses, hence the usual preference for the latter in smaller cities
- When operated in mixed traffic, trams are more likely to be delayed by disruptions in their lane. Buses, by contrast, can easily manoeuvre around obstacles. Opinions differ about whether deference that drivers show to trams — a cultural issue that varies by country — is sufficient to counteract this disadvantage.
- Tram tracks can be dangerous for cyclists, as bikes, particularly those with narrow tyres, may get their wheels caught in the track grooves. It is also possible to close the grooves of the tracks on critical sections by rubber profiles that are pressed down by the wheelflanges of the passing tram, but cannot be lowered by the weight of a cyclist. These tend not to be maintained, lessening their effectiveness over time. Crossing tracks without trouble requires a sufficient angle of crossing, reducing a cyclists' ability to avoid road hazards where tracks run along the road, especially in wet weather. This and problems with parked cars are lessened by building tracks and platforms in the middle of the road.
- Tram infrastructure occupies urban space above ground and requires modifications to traffic flow.
- Steel wheel trams are noisier than rubber-wheeled trolleybuses when cornering if there are no additional measures taken (e.g. greasing wheelflanges, which is standard in new-built systems).
- Tram drivers can control the switches ahead of them. This caused a major derailment in Geneva, Switzerland. A Wikinews article on the derailment In modern tram systems this problem has been resolved by use of switches that inhibit relocation when a tram is detected passing and/or more sophisticated means of command transmission.
- In urban areas where stops are close together, trams tend to coast between stops.
- Light rail vehicles are often heavier per passenger carried than heavy rail and monorail cars.
- The opening of new tram and light rail systems has sometimes been accompanied by a marked increase in car accidents, as a result of drivers' unfamiliarity with the physics and geometry of trolleys. Though such increases may be temporary, long-term conflicts between motorists and light rail operations can be alleviated by segregating their respective rights-of-way and installing appropriate signage and warning systems.
- Rail transport can expose neighboring populations to moderate levels of low-frequency noise. However, transportation planners use noise mitigation strategies to minimize these effects. Most of all, the potential for decreased private motor vehicle operations along the trolley's service line due to the service provision could result in lower ambient noise levels than without.

Tram tracks can be hazardous to cyclists


Tram accident in Amsterdam

On Balance

Many of the pros and cons depend on the system design itself. A tram system with little distance between stops that has single unit vehicles which run in mixed traffic will see far less of an advantage over other transit alternatives than a tram system with a greater distance between stops, runs in multiple units, and runs in a dedicated right of way. Overall trams have a greater versatility in design, however as shown above, whether that is a pro or a con is debatable.

Tram and light-rail transit systems around the world

Around the world there are many tram systems; some date back from the early 20th century but countless number of the old systems were closed down with the exception of many Eastern Europe countries in the mid-20th century. Even though many of the systems have closed down over the years there are still tram systems that have been operating much as they did when they were first built over a century ago. Some cities that have once closed down their tram networks are now in the stages of reconstructing their tramways.

Tram manufacturers

- Crotram Croatia
- Bombardier Transportation Canada
  - Urban Transportation Development Corporation 1973-1990s
  - Hawker Siddeley Canada 1962-2001
- Russell Car Company
- J. G. Brill and Company
  - American Car Company
- Ottawa Car Company
- St. Louis Car Company
- Canadian Car and Foundry
- Siemens Germany
- Ansaldobreda Italy
- Firema Transporti SpA
- Alstom France
- TDI
- CAF Spain
- Tatra Czech Republic
- Škoda Czech Republic
- Stadler Switzerland
- Dick, Kerr & Co. England
- English Electric England
- Hong Kong Tramways Hong Kong

Hong Kong Tramways passing each other at Central.

Trams in literature

One of the earliest literary references to trams occurs on the second page of Henry James's novel ***The Europeans***:

> *From time to time a strange vehicle drew near to the place where they stood - such a vehicle as the lady at the window, in spite of a considerable acquaintance with human inventions, had never seen before: a huge, low, omnibus, painted in brilliant colours, and decorated apparently with jingling*

bells, attached to a species of groove in the pavement, through which it was dragged, with a great deal of rumbling, bouncing, and scratching, by a couple of remarkably small horses.

Published in 1878, the novel is set in the 1840s, though horse trams were not in fact introduced in Boston till the 1850s. Note how the tram's efficiency surprises the "European" visitor; how two "remarkably small" horses sufficed to draw the "huge" tramcar.

Gdansk trams figure extensively in the early stages of Günter Grass's ***Die Blechtrommel*** (The Tin Drum). Then in its last chapter, the novel's hero Oskar Matzerath, along with his friend Gottfried von Vittlar, steal a tram late at night from outside the Unterrath depot on the northern edge of Düsseldorf.

It is a surreal journey. Gottfried von Vittlar drives the tram through the night, south to Flingern and Haniel and then east to the suburb of Gerresheim. Meanwhile, inside, Oskar tries to rescue the half-blind Victor Weluhn (a character who had escaped from the siege of the Polish post office in Danzig at the beginning of the book and of the war) from his two green-hatted would-be executioners. Oskar deposits his briefcase, which contains Sister Dorotea's severed ring finger in a preserving jar, on the dashboard "where professional motorman put their lunchboxes". They leave the tram at the terminus, and the executioners tie Weluhn to a tree in Vittlar's mother's garden and prepare to machine-gun him. But Oskar drums, Victor sings, and together they conjure up the Polish cavalry, who spirit both victim and executioners away. Oskar asks Vittlar to take his briefcase in the tram to the police HQ in the Fürstenwall, which he does.

The latter part of this route is today served by tram no. 703 terminating at Gerresheim Stadtbahn station ("by the glassworks" as Grass notes, referring to the famous glass factory in Gerresheim).

[Reference: The chapter *Die letzte Straßenbahn oder Anbetung eines Weckglases* (The last tram or Adoration of a Preserving Jar). See page 584 of the 1959 Büchergilde Gutenberg German edition and page 571 of the 1961 Secker & Warburg edition, translated into English by Ralph Manheim]

Trams in popular culture

- The Rev W. Awdry made a small LNER J70 tram called Toby the Tram Engine which starred in a series of books called The Railway Series along with his faithful coach, Henrietta.
- A Streetcar Named Desire (play)
- A Streetcar Named Desire (film)
- Mister Roger's Neighbourhood featured a trolley
- The central plot of the film Who Framed Roger Rabbit involves the Judge Doom, the villain, dismantling the streetcars of Los Angeles.
- "The Trolley Song" in Meet Me in St. Louis (film) received an Academy Award.
- The 1944 World Series was also known as the "Streetcar Series".
- Malcolm (film) - an Australian film about a tram enthusiast who uses his inventions to pull off a bank heist.
- Luis Bunuel filmed *La Ilusión viaja en tranvía* (English: *Illusion Travels by Stretcar*) in Mexico in 1954
- In Akira Kurosawa's film Dodesukaden a mentally ill boy pretends to be a tram conductor.
- The predominance of trams (trolleys) gave rise to the disparaging term trolley dodger for residents of the borough of Brooklyn in New York City. That

term, shortened to "Dodger" became the nickname for the Brooklyn Dodgers (now the Los Angeles Dodgers).

Transport

2008/9 Schools Wikipedia Selection. Related subjects: Engineering


Ximen Station, one of the stations of Metro Taipei.

Transport or **transportation** is the movement of people and goods from one place to another. The term is derived from the Latin *trans* ("across") and *portare* ("to carry"). Industries which have the business of providing transport equipment, transport services or transport are important in most national economies, and are referred to as **transport industries**.

The field can be divided into infrastructure, vehicles, and operations. Infrastructure consists of the fixed installations necessary for transport, and may be roads, railways, airways, waterways, canals and pipelines or terminals such as airports, railway stations, bus stations and seaports. Vehicles traveling on the network include automobiles, bicycles, buses, trains, people and aircraft. Operations deal with the way the vehicles are operated, and the procedures set for this purpose including the financing, legalities and policies.

Mode

Human-powered


Human-powered transport in front of the bulk carrier BW Fjord

Human-powered transport is the transport of person(s) and/or goods using human muscle power. Like animal-powered transport, human-powered transport has existed since time immemorial in the form of walking, running and swimming. Modern technology has allowed machines to enhance human-power. Many forms of human-powered transport remain popular for reasons of lower cost, leisure, physical exercise and environmentalism. Human-powered transport is sometimes the only type available (especially in underdeveloped or inaccessible regions), and is considered an ideal form of sustainable transportation.

Although humans are able to walk without infrastructure, the transport can be enhanced through the use of roads, especially when enforcing the human power with vehicles, such as bicycles and inline skates. Human-powered vehicles have also been developed for highly encumbering environments, ̧such as snow and water, by watercraft rowings and skiing; even the air can be entered with human-powered aircraft.


Bullock team hauling wood in Australia

Animal-powered

Animal-powered transport is the use of working animals (also known as "beasts of burden") for the movement of people and goods. Humans may ride some of the animals directly, use them as pack animals for carrying goods, or harness them, singly or in teams, to pull (or haul) sleds or wheeled vehicles. Animals are superior to people in their speed, endurance and carrying capacity; prior to the Industrial Revolution they were used for all land transport impracticable for people, and they remain an important mode of transport in less developed areas of the world.

Air


Air Canada Airbus A330 airliner

A fixed-wing aircraft, commonly called *airplane* or *aeroplane*, is a heavier-than-air craft where movement of the wings in relation to the aircraft is not used to generate lift. The term is used to distinguish from rotary-wing aircraft, where the movement of the lift surfaces relative to the aircraft generates lift. A heliplane is both fixed-wing and rotary-wing. Fixed-wing aircraft range from small trainers and recreational aircraft to large airliners and military cargo aircraft.

Two necessities for aircraft are air flow over the wings for lift, and an area for landing. The majority of aircraft also need an airport with the infrastructure to receive maintenance, restocking, refueling and for the loading and unloading of crew, cargo and passengers. While the vast majority of aircraft land and take off on land, some are capable of take off and landing on ice, snow and calm water.

The aircraft is the second fastest method of transport, after the rocket. Commercial jets can reach up to 875 km/h, single-engine aircraft 175 km/h while the speedrecond is held by the SR-71 with a speed of 3,529.56 km/h (2193.17 mph, 1905.81 knots). Aviation is able to quickly transport people and limited amounts of cargo over longer distances, but incur high costs and energy use; for short distances or in unaccessible places helicopters can be used.

Pipeline

Trans-Alaska Pipeline for crude oil.

Pipeline transport sends goods through a pipe, most commonly liquid and gases are sent, but pneumatic tubes can send solid capsules using compressed air. Any chemically stable liquid or gas can be sent through a pipeline; sewage, slurry, water and even beer pipelines exist,while long-distance networks are used for petroleum and natural gas. The idea was launched in 1863

Rail

Rail transport is the transport of passengers and goods along railways (or railroads), consisting of two parallel steel rails, generally anchored perpendicular to beams (termed sleepers or ties) of timber, concrete or steel to maintain a consistent distance apart, or gauge. The rails and perpendicular beams are usually then placed on a foundation made of concrete or compressed earth and gravel in a bed of ballast to prevent the track from buckling (bending out of its original configuration) as the ground settles over time beneath and under the weight of the vehicles passing above. The vehicles traveling on the rails are arranged in a train; a series of individual powered or unpowered vehicles linked together, displaying markers. These vehicles (referred to, in general, as cars, carriages or wagons) move with much less friction than on rubber tires on a paved road, and the locomotive that pulls the train tends to use energy far more efficiently as a result.

InterCityExpress, a German high-speed passenger train

A train consists of rail vehicles that move along guides to transport freight or passengers from one place to another. The guideway (permanent way) usually consists of conventional rail tracks, but might also be monorail or maglev. Propulsion for the train is provided by a separate locomotive, or from individual motors in self-propelled multiple units. Most trains are powered by diesel engines or by electricity supplied by trackside systems, but other sources of power such as steam engine, horses, wire, gravity, pneumatics, or gas turbines are possible.

Stone railways were constructed by the Greeks in the 6th century BC, while the first iron rails were laid in 1768 with the steam engine introduced in 1804. A critical part of industrialization, tracks were soon laid throughout the world. By the end of the century electric traction evolved, supplemented by diesel in the next. High-speed rail was introduced by Shinkansen in 1964. Rail transport remains the most energy efficient land transport, and used for long-distance freight and all distances of passenger transport. In cities rapid transit and trams are common parts of public transport.

Road

A road is an identifiable route, way or path between two or more places. Roads are typically smoothed, paved, or otherwise prepared to allow easy travel; though they need not be, and historically many roads were simply recognizable routes without any formal construction or maintenance. In urban areas roads may pass through a city or village and be named as streets, serving a dual function as urban space easement and route.

Interstate 80 near Berkeley, United States

The most common road vehicle is the automobile; a wheeled passenger vehicle that carries its own motor. Other users of roads include buses, trucks, motorcycles, bicycles and pedestrians. As of 2002 there were 590 million automobiles worldwide.

The first forms of road transport were horses, oxen or even humans carrying goods over dirt tracks that often followed game trails. The Roman Empire was in need for armies to be able to travel quickly; they built deep roadbeds of crushed stone as an underlying layer to ensure that they kept dry, as the water would flow out from the crushed stone, instead of becoming mud in clay soils. John Loudon McAdam designed the first modern highways of inexpensive paving material of soil and stone aggregate known as macadam during the Industrial Revolution. Coating of cobblestones and wooden paving were popular during the 19th century while tarmac and concrete paving became popular during the 20th.

Automobiles offer high flexibility and with low capacity, but are deemed with high energy and area use, and the main source of noise and air pollution in cities;

buses allow for more efficient travel at the cost of reduced flexibility. Road transport by truck is often the initial and final stage of freight transport.

Water

Automobile ferry in Croatia

Ship transport is the process of transport by barge, boat, ship or sailboat over a sea, ocean, lake, canal or river. A watercraft is a vehicle designed to float on and move across (or under) water. The need for buoyancy unites watercraft, and makes the hull a dominant aspect of its construction, maintenance and appearance.

The first craft were probably types of canoes cut out from tree trunks. The colonization of Australia by Indigenous Australians provides indirect but conclusive evidence for the latest date for the invention of ocean-going craft. Early sea transport was accomplished with ships that were either rowed or used the wind for propulsion, or a combination of the two. other solutions include horse-powered boats, with horses on the deck providing power.

In the 1800s the first steam ships were developed, using a steam engine to drive a paddle wheel or propeller to move the ship. The steam was produced using wood or coal. Now most ships have an engine using a slightly refined type of petroleum called bunker fuel. Some specialized ships, such as submarines, use nuclear power to produce the steam. Recreational or educational craft still use wind power, while some smaller craft use internal combustion engines to drive one or more propellers, or in the case of jet boats, an inboard water jet. In shallow draft areas hovercraft are propelled by large pusher-prop fans.

Although slow, modern sea transport is a highly effective method of transporting large quantities of non-perishable goods. Transport by water is significantly less costly than air transport for trans-continental shipping.

Intermodal transport

Freight train with containers in the United Kingdom.

Intermodal freight transport is the combination of multiple modes of transportation for a single shipment; containers allow seamless integration of sea, rail and road transport and have reduced transshipment costs.

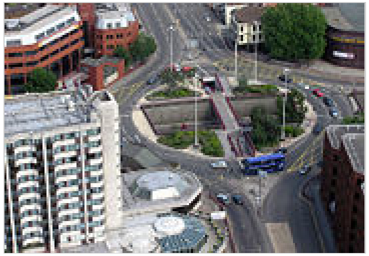
Intermodal passenger transport is where a journey is performed through the use of several modes of transport; since all human transport normally starts and ends with walking, all passenger transport can be considered intermodal. Public transport may also involve the intermediate change of vehicle, within or across modes, at at transport hub, such as a bus- or railway station.

Impact

Transport is a key necessity for specialization—allowing production and consumption of product to occur at different locations. Transport has throughout history been the gate to expansion; better transport allows more trade and spread of people. Economic growth has always been dependent on increased capacity and more rational transport. But the infrastructure and operation of transport incurs large impact on the land and is the largest drainer of energy, and transport sustainability has become a major issue.

Modern society dictates a physical distinction between home and work, forcing people to transport themselves to place of work or study, supplemented by the need to temporarily relocate for other daily activities. Passenger transport is also the essence tourism, a mayor part of recreational transport. Commerce needs transport of people to conduct business, either to allow face-to-face communication for important decisions, or to transport specialists from their regular place of work to sites where they are needed.

Planning



The engineering of this roundabout in Bristol, United Kingdom, attempts to make traffic flow free-moving.

Transport planning allows for high utilization and less impact regarding new infrastructure. Using models of transport forecasting, planners are able to predict future transport patters. On the operative level, logistics allows owners of cargo to plan transport as part of the supply chain. Transport as a field is studied through transport economics, the backbone for the creation of regulation policy by authorities.

Transport engineering, a sub-discipline of civil engineering, and must take into account trip generation, trip distribution, mode choice and route assignment, while the operative level is handles through traffic engineering.

Because of the negative impacts made, transport often becomes the subject of controversy related to choice of mode, as well as increased capacity. Automotive transport can be seen as a tragedy of the commons, where the flexibility and comfort for the individual deteriorate the natural and urban environment for all. Density of development depends on mode of transport, with public transport allowing for better spacial utilization. Good land use keeps common activities close to peoples homes and places higher-density development closer to transport lines and hubs; minimize the need for transport. There are economies of agglomeration. Beyond transportation some land uses are more efficient when clustered. Transportation facilities consume land, and in cities, pavement (devoted to streets and parking) can easily exceed 20 percent of the total land use. An efficient transport system can reduce land waste.

Too much infrastructure and too much smoothing for maximum vehicle throughput means that in many cities there is too much traffic and many—if not all—of the negative impacts that come with it. It is only in recent years that traditional practices have started to be questioned in many places, and as a result of new types of analysis which bring in a much broader range of skills than those traditionally relied on—spanning such areas as environmental impact analysis, public health, sociologists as well as economists who increasingly are questioning the viability of the old mobility solutions. European cities are leading this transition.

Construction

The financing of infrastructure can either be public or private; since it often is a natural monopoly and a necessity for the public, roads, and in some countries railways and airports are funded through taxation. New infrastructure projects can involve large spendings, and encompass many megaprojects.

Traffic congestion persists in São Paulo, Brasil despite of the no-drive days based on license numbers.

Operations may be public, but airlines and road transport is commonly private, with the typical exception of mass transit. Shipping remains a highly competitive industry with little regulation, but ports can be public owned.

Environment

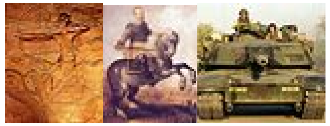
Transport is a major use of energy, burning most of the world's petroleum; creating air pollution, including nitrous oxides and particulates and being a significant contributor to global warming through emission of carbon dioxide—the fastest growing emission sector. Environmental regulations in developed countries have reduced the individual vehicles emission; this has been offset by an increase in the number of vehicles and more use of each vehicle. Energy use and emissions vary largely between modes, causing |environmentalists to call for a transition from air and road to rail, sea and human-powered transport. Other environmental impacts of transport systems include traffic congestion and automobile-oriented urban sprawl, which can consume natural habitat and agricultural lands.

Weapon

2008/9 Schools Wikipedia Selection. Related subjects: Engineering; Military History and War

A **weapon** is a tool employed to gain a tactical advantage over an adversary, usually by injury, defeat, or destruction.. There are a huge variety of weapons, which all have different means of coercion. Weapons may be used to attack and defend, and consequently also to threaten or protect. Metaphorically, anything used to damage (even psychologically) can be referred to as a weapon. A weapon can be as simple as a club or as complex as an intercontinental ballistic missile.

War



Military History

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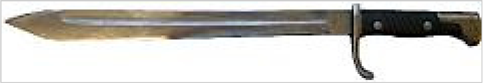
Strategy	[Show]
Organization	[Show]
Logistics	[Show]
Lists	[Show]

War Portal

History



The bayonet is used as both knife and spear.

A weapon is an object that is used to increase the destructive range and/or power of the wielder. Weapons are used in the hunting of animals, as well as in conflicts between humans. From the earliest traces of mankind up to modern civilization, weapons have been a facet of human development. In modern times, weapons development has accelerated along with technological development in general. In ancient times, from the dawn of humanity, through the Classical civilizations of Greece and Rome, and until the widespread introduction of gunpowder weaponry in the Renaissance, weapons were primarily extensions of an individual's strength, essentially making up for the human body's lack of natural weapons such as claws. These weapons allowed the bearer to be substantially more deadly and lethal than a similar human without such a weapon. Although many weapons made in ancient times were steel, wooden ones were also very common.

The earliest weapons used by man for hunting purposes are the Schöninger Speere, eight wooden spears discovered between 1995 and 1998 in a surface mine in Schöningen, Germany. According to various sources, they are about 400,000 years old, making them part of human life in the Lower Paleolithic era.

The Medieval period, including the Western Middle Ages, was characterized by two iconic weapons: knights, heavily-armored horsemen, and castles, fortified dwellings which proliferated throughout Europe and the near est. While knights harked back to earlier historical cavalry such as the Roman and Persian cataphracts, castles triggered quite revolutionary advances, including increasingly sophisticated siegecraft.

The Renaissance marked the beginning of the implementation of combustion powered devices in warfare. The most long-lasting effect of this was the introduction of guns and rockets to the battlefield, which are still at the core of modern weaponry. However, many other machines of war were experimented with.

From the American Revolution through the beginning of the 20th century, human-powered weapons were finally excluded from the battlefield for the most part. Sometimes referred to as the "Age of Rifles", this period was characterized by the development of firearms for infantry and cannons for support, as well as the beginnings of mechanized weapons such as the machine gun.

Ancient Chinese cannon displayed in the Tower of London.

World War I marked the entry of fully industrialized warfare, and weapons were developed quickly to meet wartime needs. Many new technologies were developed, particularly in the development of military aircraft and vehicles. World War II however, perhaps marked the most frantic period of weapons development in the history of humanity. Massive numbers of new designs and concepts were fielded, and all existing technologies were improved between 1939 and 1945. Ultimately, the most powerful of all invented weapons was the Hydrogen bomb.

After World War II, with the onset of the Cold War, the constant technological development of new weapons was institutionalized, as participants engaged in a constant race to develop weapons and counter-weapons. This constant state of weapons development continues into the modern era, and remains a constant draw on the resources of most nations.

Combustion-powered weapons

Firearms are qualitatively different from earlier weapons because they store energy in a combustible propellant such as gunpowder, rather than in a weight or spring. This energy is released quite rapidly, and can be restored without much effort by the user, so that even early firearms such as the arquebus were much more powerful than human-powered weapons. They became increasingly important and effective during the 16th century to 19th century, with progressive improvements in ignition mechanisms followed by revolutionary changes in ammunition handling and propellant. During the U.S. Civil War various technologies including the machine gun and ironclad warship emerged that would be recognizable and useful military weapons today, particularly in lower-technology conflicts. In the 19th century warship propulsion changed from sail power to fossil fuel-powered steam engines.

The Maxim gun and its derivative the Vickers (shown here) remained in British military service for 79 consecutive years.

The age of edged weapons ended abruptly just before World War I with rifled artillery, such as howitzers which are able to destroy any masonry fortress. This single invention caused a revolution in military affairs and doctrines that continues to this day. *See Technology during World War I for a detailed discussion.*

An important feature of industrial age warfare was technological escalation - an innovation could, and would, be rapidly matched by copying it, and often with yet another innovation to counter it. The technological escalation during World War I was profound, producing armed aircraft and tanks.

This continued in the period between the end of that war and the next, with continuous improvements of all weapons by all major powers. Many modern military weapons, particularly ground-based ones, are relatively minor improvements on those of World War II. *See military technology during World War II for a detailed discussion.*

The most notable development in weaponry since World War II has been the combination and further development of two weapons first used in it—nuclear weapons and the ballistic missile, leading to its ultimate configuration: the ICBM. The mutual possession of these by the United States and the Soviet Union ensured that either nation could inflict terrible damage on the other; so terrible, in fact, that neither nation was willing to instigate direct, all-out war with the other (a phenomenon known as Mutually Assured Destruction). The indiscriminate nature of the destruction has made nuclear-tipped missiles essentially useless for the smaller wars fought since. However computer-guided weaponry of all kinds, from precision-guided munitions (or "smart bombs") to computer-aimed tank rounds, has greatly increased weaponry's accuracy.

India's Agni-II, a ballistic missile. (Photo: Antônio Milena/ABr)

Information warfare

In modern warfare, since all redoubts are traps, maneuver and coordination of forces are decisive, overshadowing particular weapons. The goal of every modern commander is therefore to "operate within the observation-decision-action cycle of the enemy." In this way, the modern commander can bring overwhelming force to bear on isolated groups of the enemy, and "tactically" overwhelm an enemy.

Traditional military maneuvers tried to achieve this coordination with "fronts" made of lines of military assets. These were formerly the only way to prevent harm to friendly forces. Close-order marching and drill (a traditional military skill) was an early method to get relative superiority of coordination. Derivative methods (such as "leapfrogging units to advance a line") survived into combined arms warfare to coordinate aircraft, artillery, armor and infantry.

Computers are changing this. Attacks are throroughly navigated with great precision.

Thus in modern warfare, satellite navigation systems, digital radios and computers give decisive advantages to ordinary military personnel armed with weapons that are otherwise unremarkable.

Weapon types

There are essentially three facets to classifying weapon types: who uses it, how it works, and what it targets.

Who uses it essentially determines how it can be employed:

- Personal weapons (or Small Arms) are designed to be used by a single person.
- Crew served weapons are larger than personal weapons, requiring more than one crew member to operate correctly.

- Fortification weapons are designed to be mounted in a permanent installation, or used primarily within a fortification.
- Mountain weapons are designed for use by mountain forces or those operating in difficult terrain and harsh climates.
- Vehicle weapons are designed to be mounted on any type of military vehicle.
- Railway weapons are designed to be mounted on railway cars, including armored trains.
- Aircraft weapons are designed to be carried on and used by some type of aircraft, helicopter, or other aerial vehicle.
- Naval weapons are designed to be mounted on ships and submarines.
- Space weapons are designed to be used in or launched from space.

How it works refers to the construction of the weapon and how it operates:

- Antimatter weapons (still theoretical) would combine matter and antimatter to cause a powerful explosion. However, antimatter is still hard to make and harder to store.
- Archery related weapons operate by using a tensioned string to launch a projectile at some target.
- Artillery are large firearms capable of launching heavy projectiles (normally explosive) over long distances.
- Biological weapons spread biological agents, attacking humans (or livestock) by causing disease and infection.
- Chemical weapons spread chemical agents, attacking humans by poisoning and causing reactions.
- Energy weapons rely on concentrating forms of energy to attack, such as lasers, electrical shocks, and thermal or sonic attack.
- Explosive weapons use a physical explosion to create blast concussion or spread shrapnel.
- Firearms use a chemical charge to launch one or more projectiles down a rifled or smoothbore barrel.
- Future weapons make use of futuristic high-tech weapon systems and advanced materials.
- Improvised weapons are common objects that were not designed for combat purposes but are used as such in self defense, guerrilla warfare or a violent crime.
- Incendiary weapons rely on combustible materials and an ignition mechanism to cause damage by fire.
- Non-lethal weapons are used to attack and subdue humans, but are designed to minimize the risk of killing the target.
- Magnetic weapons is one that uses magnetic fields to accelerate and propel projectiles, or to focus charged particle beams.
- Mêlée weapons operate as physical extensions of the user's body and directly impact their target.
- Missiles are rockets which are guided to their target after launch. This is also a general term for projectile weapons.
- Nuclear weapons use radioactive materials to create nuclear fission and/or nuclear fusion detonations above a target ("air-burst") or at ground-level.
- Primitive weapons make little or no use of technological or industrial elements, instead being purely constructed of easily obtainable natural materials.
- Ranged weapons cause a projectile to leave the user and (ideally) strike a target afterwards.
- Rockets use chemical propellant to accelerate a projectile (usually with an explosive warhead) towards a target and are typically unguided once fired.
- Suicide weapons are typically explosive in nature and exploit the willingness of their operator to not survive the attack to reach their target.

What it targets refers to what type of target the weapon is designed to attack:

- Anti-aircraft weapons target enemy aircraft, helicopters, missiles and any other aerial vehicles in flight.
- Anti-fortification weapons are designed to target enemy installations, including bunkers and fortifications. The American bunker buster bomb is designed

to travel almost 10 metres underground before detonating, toppling underground installations.
- Anti-personnel weapons are designed to attack people, either individually or in numbers.
- Anti-radiation weapons target enemy sources of electronic radiation, particularly radar emitters.
- Anti-ship weapons target enemy ships and vessels on water.
- Anti-submarine weapons target enemy submarines and other underwater targets.
- Anti-tank weapons are primarily used to defeat armored targets, but may be targeted against other less well armored targets.
- Area denial weapons are designed to target territory, making it unsafe or unsuitable for enemy use or travel.
- Hunting weapons are designed particularly for use against animals for hunting purposes.
- Infantry support weapons are designed to attack various threats to infantry units, supporting the infantry's operations, including heavy machine guns, mortars and pinpoint airstrikes ordered by the infantry, often to strike heavily defended positions, such as enemy camps or extensively powerful machine-gun nests.

Weapons by era

- Ancient
- Medieval
- Technology during World War I
- Military technology during World War II
- Modern weapons
    - All eras

Welding

2008/9 Schools Wikipedia Selection. Related subjects: Engineering

Welding is a fabrication process that joins materials, usually metals or thermoplastics, by causing coalescence. This is often done by melting the workpieces and adding a filler material to form a pool of molten material (the *weld puddle*) that cools to become a strong joint, with pressure sometimes used in conjunction with heat, or by itself, to produce the weld. This is in contrast with soldering and brazing, which involve melting a lower-melting-point material between the workpieces to form a bond between them, without melting the workpieces.


Arc welding

Many different energy sources can be used for welding, including a gas flame, an electric arc, a laser, an electron beam, friction, and ultrasound. While often an industrial process, welding can be done in many different environments, including open air, underwater and in space. Regardless of location, however, welding remains dangerous, and precautions must be taken to avoid burns, electric shock, eye damage, poisonous fumes, and overexposure to ultraviolet light.

Until the end of the 19th century, the only welding process was forge welding, which blacksmiths had used for centuries to join metals by heating and pounding them. Arc welding and oxyfuel welding were among the first processes to develop late in the century, and resistance welding followed soon after. Welding technology advanced quickly during the early 20th century as World War I and World War II drove the demand for reliable and inexpensive joining methods. Following the wars, several modern welding techniques were developed, including manual methods like shielded metal arc welding, now one of the most popular welding methods, as well as semi-automatic and automatic processes such as gas metal arc welding, submerged arc welding, flux-cored arc welding and electroslag welding. Developments continued with the invention of laser beam welding and electron beam welding in the latter half of the century. Today, the science continues to advance. Robot welding is becoming more commonplace in industrial settings, and researchers continue to develop new welding methods and gain greater understanding of weld quality and properties.

History


The Iron pillar of Delhi.

The history of joining metals goes back several millennia, with the earliest examples of welding from the Bronze Age and the Iron Age in Europe and the Middle East. Welding was used in the construction of the Iron pillar in Delhi, India, erected about 310 and weighing 5.4 metric tons. The Middle Ages brought advances in forge welding, in which blacksmiths pounded heated metal repeatedly until bonding occurred. In 1540, Vannoccio Biringuccio published *De la pirotechnia*, which includes descriptions of the forging operation. Renaissance craftsmen were skilled in the process, and the industry continued to grow during the following centuries. Welding, however, was transformed during the 19th century—in 1800, Sir Humphry Davy discovered the electric arc, and advances in arc welding continued with the inventions of metal electrodes by a Russian, Nikolai Slavyanov, and an American, C. L. Coffin in the late 1800s, even as carbon arc welding, which used a carbon electrode, gained popularity. Around 1900, A. P. Strohmenger released a coated metal electrode in Britain, which gave a more stable arc, and in 1919, alternating current welding was invented by C. J. Holslag, but did not become popular for another decade.

Resistance welding was also developed during the final decades of the 19th century, with the first patents going to Elihu Thomson in 1885, who produced further advances over the next 15 years. Thermite welding was invented in 1893, and around that time, another process, oxyfuel welding, became well established. Acetylene was discovered in 1836 by Edmund Davy, but its use was not practical in welding until about 1900, when a suitable blowtorch was developed. At first, oxyfuel welding was one of the more popular welding methods due to its portability and relatively low cost. As the 20th century progressed, however, it fell out of favour for industrial applications. It was largely replaced with arc welding, as metal coverings (known as flux) for the electrode that stabilize the arc and shield the base material from impurities continued to be developed.

World War I caused a major surge in the use of welding processes, with the various military powers attempting to determine which of the several new welding processes would be best. The British primarily used arc welding, even constructing a ship, the *Fulagar*, with an entirely welded hull. The Americans were more hesitant, but began to recognize the benefits of arc welding when the process allowed them to repair their ships quickly after German attacks in the New York Harbour at the beginning of the war. Arc welding was first applied to aircraft during the war as well, as some German airplane fuselages were constructed using the process.. Also noteworthy is the first welded road bridge in the world built across the river Słudwia Maurzyce near Łowicz, Poland) in 1929, but designed by Stefan Bryła of the Warsaw University of Technology in 1927.

During the 1920s, major advances were made in welding technology, including the introduction of automatic welding in 1920, in which electrode wire was fed continuously. Shielding gas became a subject receiving much attention, as scientists attempted to protect welds from the effects of oxygen and nitrogen in the atmosphere. Porosity and brittleness were the primary problems, and the solutions that developed included the use of hydrogen, argon, and helium as welding atmospheres. During the following decade, further advances allowed for the welding of reactive metals like aluminium and magnesium. This, in conjunction with developments in automatic welding, alternating current, and fluxes fed a major expansion of arc welding during the 1930s and then during World War II.

During the middle of the century, many new welding methods were invented. 1930 saw the release of stud welding, which soon became popular in shipbuilding and construction. Submerged arc welding was invented the same year, and continues to be popular today. Gas tungsten arc welding, after decades of development, was finally perfected in 1941, and gas metal arc welding followed in 1948, allowing for fast welding of non- ferrous materials but requiring expensive shielding gases. Shielded metal arc welding was developed during the 1950s, using a flux coated consumable electrode, and it quickly became the most popular metal arc welding process. In 1957, the flux-cored arc welding process debuted, in which the self-shielded wire electrode could be used with automatic equipment, resulting in greatly increased welding speeds, and that same year, plasma arc welding was invented. Electroslag welding was introduced in 1958, and it was followed by its cousin, electrogas welding, in 1961.

Other recent developments in welding include the 1958 breakthrough of electron beam welding, making deep and narrow welding possible through the concentrated heat source. Following the invention of the laser in 1960, laser beam welding debuted several decades later, and has proved to be especially useful in high-speed, automated welding. Both of these processes, however, continue to be quite expensive due the high cost of the necessary equipment, and this has limited their applications.

Welding processes

Arc welding

These processes use a welding power supply to create and maintain an electric arc between an electrode and the base material to melt metals at the welding point. They can use either direct (DC) or alternating (AC) current, and consumable or non-consumable electrodes. The welding region is sometimes protected by some type of inert or semi- inert gas, known as a shielding gas, and filler material is sometimes used as well.

Power supplies

To supply the electrical energy necessary for arc welding processes, a number of different power supplies can be used. The most common classification is constant current power supplies and constant voltage power supplies. In arc welding, the length of the arc is directly related to the voltage, and the amount of heat input is related to the current. Constant current power supplies are most often used for manual welding processes such as gas tungsten arc welding and shielded metal arc welding, because they maintain a relatively constant current even as the voltage varies. This is important because in manual welding, it can be difficult to hold the electrode perfectly steady, and as a result, the arc length and thus voltage tend to fluctuate. Constant voltage power supplies hold the voltage constant and vary the current, and as a result, are most often used for automated welding processes such as gas metal arc welding, flux cored arc welding, and submerged arc welding. In these processes, arc length is kept constant, since any fluctuation in the distance between the wire and the base material is quickly rectified by a large change in current. For example, if the wire and the base material get too close, the current will rapidly increase, which in turn causes the heat to increase and the tip of the wire to melt, returning it to its original separation distance.

The type of current used in arc welding also plays an important role in welding. Consumable electrode processes such as shielded metal arc welding and gas metal arc welding generally use direct current, but the electrode can be charged either positively or negatively. In welding, the positively charged anode will have a greater heat concentration, and as a result, changing the polarity of the electrode has an impact on weld properties. If the electrode is positively charged, the base metal will be hotter, increasing weld penetration and welding speed. Alternatively, a negatively charged electrode results in more shallow welds. Nonconsumable electrode processes, such as gas tungsten arc welding, can use either type of direct current, as well as alternating current. However, with direct current, because the electrode only creates the arc and does not provide filler material, a positively charged electrode causes shallow welds, while a negatively charged electrode makes deeper welds. Alternating current rapidly moves between these two, resulting in medium-penetration welds. One disadvantage of AC, the fact that the arc must be re-ignited after every zero crossing, has been addressed with the invention of special power units that produce a square wave pattern instead of the normal sine wave, making rapid zero crossings possible and minimizing the effects of the problem.

Processes


Shielded metal arc welding

One of the most common types of arc welding is shielded metal arc welding (SMAW), which is also known as manual metal arc welding (MMA) or stick welding. Electric current is used to strike an arc between the base material and consumable electrode rod, which is made of steel and is covered with a flux that protects the weld area from oxidation and contamination by producing CO_2 gas during the welding process. The electrode core itself acts as filler material, making a separate filler unnecessary.

The process is versatile and can be performed with relatively inexpensive equipment, making it well suited to shop jobs and field work. An operator can become reasonably proficient with a modest amount of training and can achieve mastery with experience. Weld times are rather slow, since the consumable electrodes must be frequently replaced and because slag, the residue from the flux, must be chipped away after welding. Furthermore, the process is generally limited to welding ferrous materials, though special electrodes have made possible the welding of cast iron, nickel, aluminium, copper, and other metals. Inexperienced operators may find it difficult to make good out-of-position welds with this process.

Gas metal arc welding (GMAW), also known as metal inert gas or MIG welding, is a semi-automatic or automatic process that uses a continuous wire feed as an electrode and an inert or semi-inert gas mixture to protect the weld from contamination. As with SMAW, reasonable operator proficiency can be achieved with modest training. Since the electrode is continuous, welding speeds are greater for GMAW than for SMAW. Also, the smaller arc size compared to the shielded metal arc welding process makes it easier to make out-of-position welds (e.g., overhead joints, as would be welded underneath a structure).

The equipment required to perform the GMAW process is more complex and expensive than that required for SMAW, and requires a more complex setup procedure. Therefore, GMAW is less portable and versatile, and due to the use of a separate shielding gas, is not particularly suitable for outdoor work. However, owing to the higher average rate at which welds can be completed, GMAW is well suited to production welding. The process can be applied to a wide variety of metals, both ferrous and non-ferrous.

A related process, flux-cored arc welding (FCAW), uses similar equipment but uses wire consisting of a steel electrode surrounding a powder fill material. This cored wire is more expensive than the standard solid wire and can generate fumes and/or slag, but it permits even higher welding speed and greater metal penetration.

Gas tungsten arc welding (GTAW), or tungsten inert gas (TIG) welding (also sometimes erroneously referred to as heliarc welding), is a manual welding process that uses a nonconsumable tungsten electrode, an inert or semi-inert gas mixture, and a separate filler material. Especially useful for welding thin materials, this method is characterized by a stable arc and high quality welds, but it requires significant operator skill and can only be accomplished at relatively low speeds.

GTAW can be used on nearly all weldable metals, though it is most often applied to stainless steel and light metals. It is often used when quality welds are extremely important, such as in bicycle, aircraft and naval applications. A related process, plasma arc welding, also uses a tungsten electrode but uses plasma gas to make the arc. The arc is more concentrated than the GTAW arc, making transverse control more critical and thus generally restricting the technique to a mechanized process. Because of its stable current, the method can be used on a wider range of material thicknesses than can the GTAW process, and furthermore, it is much faster. It can be applied to all of the same materials as GTAW except magnesium, and automated welding of stainless steel is one important application of the process. A variation of the process is plasma cutting, an efficient steel cutting process.

Submerged arc welding (SAW) is a high-productivity welding method in which the arc is struck beneath a covering layer of flux. This increases arc quality, since contaminants in the atmosphere are blocked by the flux. The slag that forms on the weld generally comes off by itself, and combined with the use of a continuous wire feed, the weld deposition rate is high. Working conditions are much improved over other arc welding processes, since the flux hides the arc and almost no smoke is produced. The process is commonly used in industry, especially for large products and in the manufacture of welded pressure vessels. Other arc welding processes include atomic hydrogen welding, carbon arc welding, electroslag welding, electrogas welding, and stud arc welding.

Gas welding a steel armature using the oxy-acetylene process.

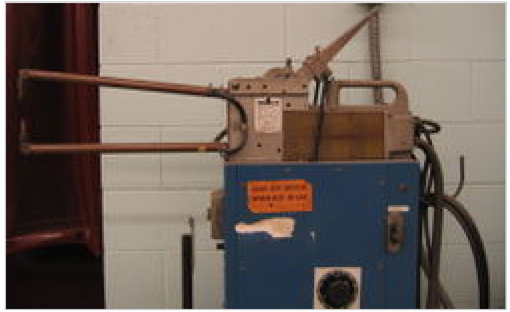
Gas welding

The most common gas welding process is oxyfuel welding, also known as oxyacetylene welding. It is one of the oldest and most versatile welding processes, but in recent years it has become less popular in industrial applications. It is still widely used for welding pipes and tubes, as well as repair work. It is also frequently well-suited, and favored, for fabricating some types of metal-based artwork. Oxyfuel equipment is versatile, lending itself not only to some sorts of iron or steel welding but also to brazing, braze-welding, metal heating (for bending and forming), and also oxyfuel cutting.

The equipment is relatively inexpensive and simple, generally employing the combustion of acetylene in oxygen to produce a welding flame temperature of about 3100 °C. The flame, since it is less concentrated than an electric arc, causes slower weld cooling, which can lead to greater residual stresses and weld distortion, though it eases the welding of high alloy steels. A similar process, generally called oxyfuel cutting, is used to cut metals. Other gas welding methods, such as air acetylene welding, oxygen hydrogen welding, and pressure gas welding are quite similar, generally differing only in the type of gases used. A water torch is sometimes used for precision welding of small items such as jewelry. Gas welding is also used in plastic welding, though the heated substance is air, and the temperatures are much lower.

Resistance welding

Resistance welding involves the generation of heat by passing current through the resistance caused by the contact between two or more metal surfaces. Small pools of molten metal are formed at the weld area as high current (1000–100,000 A) is passed through the metal. In general, resistance welding methods are efficient and cause little pollution, but their applications are somewhat limited and the equipment cost can be high.


Spot welder

Spot welding is a popular resistance welding method used to join overlapping metal sheets of up to 3 mm thick. Two electrodes are simultaneously used to clamp the metal sheets together and to pass current through the sheets. The advantages of the method include efficient energy use, limited workpiece deformation, high production rates, easy automation, and no required filler materials. Weld strength is significantly lower than with other welding methods, making the process suitable for only certain applications. It is used extensively in the automotive industry—ordinary cars can have several thousand spot welds made by industrial robots. A specialized process, called shot welding, can be used to spot weld stainless steel.

Like spot welding, seam welding relies on two electrodes to apply pressure and current to join metal sheets. However, instead of pointed electrodes, wheel-shaped electrodes roll along and often feed the workpiece, making it possible to make long continuous welds. In the past, this process was used in the manufacture of beverage cans, but now its uses are more limited. Other resistance welding methods include flash welding, projection welding, and upset welding.

Energy beam welding

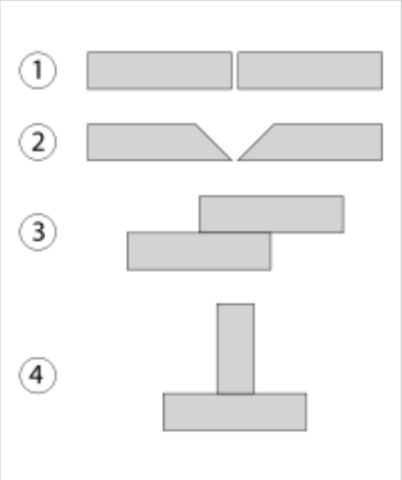
Energy beam welding methods, namely laser beam welding and electron beam welding, are relatively new processes that have become quite popular in high production applications. The two processes are quite similar, differing most notably in their source of power. Laser beam welding employs a highly focused laser beam, while electron beam welding is done in a vacuum and uses an electron beam. Both have a very high energy density, making deep weld penetration possible and minimizing the size of the weld area. Both processes are extremely fast, and are easily automated, making them highly productive. The primary disadvantages are their very high equipment costs (though these are decreasing) and a susceptibility to thermal cracking. Developments in this area include laser-hybrid welding, which uses principles from both laser beam welding and arc welding for even better weld properties.

Solid-state welding

Like the first welding process, forge welding, some modern welding methods do not involve the melting of the materials being joined. One of the most popular, ultrasonic welding, is used to connect thin sheets or wires made of metal or thermoplastic by vibrating them at high frequency and under high pressure. The equipment and methods involved are similar to that of resistance welding, but instead of electric current, vibration provides energy input. Welding metals with this process does not involve melting the materials; instead, the weld is formed by introducing mechanical vibrations horizontally under pressure. When welding plastics, the materials should have similar melting temperatures, and the vibrations are introduced vertically. Ultrasonic welding is commonly used for making electrical connections out of aluminium or copper, and it is also a very common polymer welding process.

Another common process, explosion welding, involves the joining of materials by pushing them together under extremely high pressure. The energy from the impact plasticizes the materials, forming a weld, even though only a limited amount of heat is generated. The process is commonly used for welding dissimilar materials, such as the welding of aluminium with steel in ship hulls or compound plates. Other solid-state welding processes include co-extrusion welding, cold welding, diffusion welding, friction welding (including friction stir welding), high frequency welding, hot pressure welding, induction welding, and roll welding.

Geometry



Common welding joint types – (1) Square butt joint, (2) Single-V preparation joint, (3) Lap joint, (4) T-joint.

Welds can be geometrically prepared in many different ways. The five basic types of weld joints are the butt joint, lap joint, corner joint, edge joint, and T-joint. Other variations exist as well—for example, double-V preparation joints are characterized by the two pieces of material each tapering to a single centre point at one-half their height. Single-U and double-U preparation joints are also fairly common—instead of having straight edges like the single-V and double-V preparation joints, they are curved, forming the shape of a U. Lap joints are also commonly more than two pieces thick—depending on the process used and the thickness of the material, many pieces can be welded together in a lap joint geometry.

Often, particular joint designs are used exclusively or almost exclusively by certain welding processes. For example, resistance spot welding, laser beam welding, and electron beam welding are most frequently performed on lap joints. However, some welding methods, like shielded metal arc welding, are extremely versatile and can weld virtually any type of joint. Additionally, some processes can be used to make multipass welds, in which one weld is allowed to cool, and then another weld is performed on top of it. This allows for the welding of thick sections arranged in a single-V preparation joint, for example.


The cross-section of a welded butt joint, with the darkest gray representing the weld or fusion zone, the medium gray the heat-affected zone, and the lightest gray the base material.

After welding, a number of distinct regions can be identified in the weld area. The weld itself is called the fusion zone—more specifically, it is where the filler metal was laid during the welding process. The properties of the fusion zone depend primarily on the filler metal used, and its compatibility with the base materials. It is surrounded by the heat-affected zone, the area that had its microstructure and properties altered by the weld. These properties depend on the base material's behaviour when subjected to heat. The metal in this area is often weaker than both the base material and the fusion zone, and is also where residual stresses are found.

Quality

Most often, the major metric used for judging the quality of a weld is its strength and the strength of the material around it. Many distinct factors influence this, including the welding method, the amount and concentration of energy input, the base material, the filler material, the flux material, the design of the joint, and the interactions between all these factors. To test the quality of a weld, either destructive or nondestructive testing methods are commonly used to verify that welds are defect-free, have acceptable levels of residual stresses and distortion, and have acceptable heat-affected zone (HAZ) properties. Welding codes and specifications exist to guide welders in proper welding technique and in how to judge the quality of welds.

Heat-affected zone


The blue area results from oxidation at a corresponding temperature of 600 °F (316 °C). This is an accurate way to identify temperature, but does not represent the HAZ width. The HAZ is the narrow area that immediately surrounds the welded base metal.

The effects of welding on the material surrounding the weld can be detrimental—depending on the materials used and the heat input of the welding process used, the HAZ can be of varying size and strength. The thermal diffusivity of the base material plays a large role—if the diffusivity is high, the material cooling rate is high and the HAZ is relatively small. Conversely, a low diffusivity leads to slower cooling and a larger HAZ. The amount of heat injected by the welding process plays an important role as well, as processes like oxyacetylene welding have an unconcentrated heat input and increase the size of the HAZ. Processes like laser beam welding give a highly concentrated, limited amount of heat, resulting in a small HAZ. Arc welding falls between these two extremes, with the individual processes varying somewhat in heat input. To calculate the heat input for arc welding procedures, the following formula can be used:

$$Q = \left(\frac{V \times I \times 60}{S \times 1000}\right) \times Efficiency$$

where Q = heat input (kJ/ mm), V = voltage (V), I = current (A), and S = welding speed (mm/min). The efficiency is dependent on the welding process used, with shielded metal arc welding having a value of 0.75, gas metal arc welding and submerged arc welding, 0.9, and gas tungsten arc welding, 0.8.

Distortion and cracking

Welding methods that involve the melting of metal at the site of the joint necessarily are prone to shrinkage as the heated metal cools. Shrinkage, in turn, can introduce residual stresses and both longitudinal and rotational distortion. Distortion can pose a major problem, since the final product is not the desired shape. To alleviate rotational distortion, the workpieces can be offset, so that the welding results in a correctly shaped piece. Other methods of limiting distortion, such as clamping the workpieces in place, cause the buildup of residual stress in the heat-affected zone of the base material. These stresses can reduce the strength of the base material, and can lead to catastrophic failure through cold cracking, as in the case of several of the Liberty ships. Cold cracking is limited to steels, and is associated with the formation of martensite as the weld cools. The cracking occurs in the heat-affected zone of the base material. To reduce the amount of distortion and residual stresses, the amount of heat input should be limited, and the welding sequence used should not be from one end directly to the other, but rather in segments. The other type of cracking, hot cracking or solidification cracking, can occur with all metals, and happens in the fusion zone of a weld. To diminish the probability of this type of cracking, excess material restraint should be avoided, and a proper filler material should be utilized.

Weldability

The quality of a weld is also dependent on the combination of materials used for the base material and the filler material. Not all metals are suitable for welding, and not all filler metals work well with acceptable base materials.

Steels

The weldability of steels is inversely proportional to a property known as the hardenability of the steel, which measures the probability of forming martensite during welding or heat treatment. The hardenability of steel depends on its chemical composition, with greater quantities of carbon and other alloying elements resulting in a higher hardenability and thus a lower weldability. In order to be able to judge alloys made up of many distinct materials, a measure known as the equivalent carbon content is used to compare the relative weldabilities of different alloys by comparing their properties to a plain carbon steel. The effect on weldability of elements like chromium and vanadium, while not as great as carbon, is more significant than that of copper and nickel, for example. As the equivalent carbon content rises, the weldability of the alloy decreases. The disadvantage to using plain carbon and low-alloy steels is their lower strength—there is a trade-off between material strength and weldability. High strength, low-alloy steels were developed especially for welding applications during the 1970s, and these generally easy to weld materials have good strength, making them ideal for many welding applications.

Stainless steels, because of their high chromium content, tend to behave differently with respect to weldability than other steels. Austenitic grades of stainless steels tend to be the most weldable, but they are especially susceptible to distortion due to their high coefficient of thermal expansion. Some alloys of this type are prone to cracking and reduced corrosion resistance as well. Hot cracking is possible if the amount of ferrite in the weld is not controlled—to alleviate the problem, an electrode is used that deposits a weld metal containing a small amount of ferrite. Other types of stainless steels, such as ferritic and martensitic stainless steels, are not as easily welded, and must often be preheated and welded with special electrodes.

Aluminium

The weldability of aluminium alloys varies significantly, depending on the chemical composition of the alloy used. Aluminum alloys are susceptible to hot cracking, and to combat the problem, welders increase the welding speed to lower the heat input. Preheating reduces the temperature gradient across the weld zone and thus helps reduce hot cracking, but it can reduce the mechanical properties of the base material and should not be used when the base material is restrained. The design of the joint can be changed as well, and a more compatible filler alloy can be selected to decrease the likelihood of hot cracking. Aluminium alloys should also be cleaned prior to welding, with the goal of removing all oxides, oils, and loose particles from the surface to be welded. This is especially important because of an aluminium weld's susceptibility to porosity due to hydrogen and dross due to oxygen.

Unusual conditions

Underwater welding

While many welding applications are done in controlled environments such as factories and repair shops, some welding processes are commonly used in a wide variety of conditions, such as open air, underwater, and vacuums (such as space). In open-air applications, such as construction and outdoors repair, shielded metal arc welding is the most common process. Processes that employ inert gases to protect the weld cannot be readily used in such situations, because unpredictable atmospheric movements can result in a faulty weld. Shielded metal arc welding is also often used in underwater welding in the construction and repair of ships, offshore platforms, and pipelines, but others, such as flux cored arc welding and gas tungsten arc welding, are also common. Welding in space is also possible—it was first attempted in 1969 by Russian cosmonauts, when they performed experiments to test shielded metal arc welding, plasma arc welding, and electron beam welding in a depressurized environment. Further testing of these methods was done in the following decades, and today researchers continue to develop methods for using other welding processes in space, such as laser beam welding, resistance welding, and friction welding. Advances in these areas could prove indispensable for projects like the construction of the International Space Station, which will likely rely heavily on welding for joining in space the parts that were manufactured on Earth.

Safety issues

Arc welding with a welding helmet, gloves, and other protective clothing.

Welding, without the proper precautions, can be a dangerous and unhealthy practice. However, with the use of new technology and proper protection, risks of injury and death associated with welding can be greatly reduced. Because many common welding procedures involve an open electric arc or flame, the risk of burns is significant. To prevent them, welders wear personal protective equipment in the form of heavy leather gloves and protective long sleeve jackets to avoid exposure to extreme heat and flames. Additionally, the brightness of the weld area leads to a condition called arc eye in which ultraviolet light causes inflammation of the cornea and can burn the retinas of the eyes. Goggles and welding helmets with dark face plates are worn to prevent this exposure, and in recent years, new helmet models have been produced that feature a face plate that self-darkens upon exposure to high amounts of UV light. To protect bystanders, translucent welding curtains often surround the welding area. These curtains, made of a polyvinyl chloride plastic film, shield nearby workers from exposure to the UV light from the electric arc, but should not be used to replace the filter glass used in helmets.

Welders are also often exposed to dangerous gases and particulate matter. Processes like flux-cored arc welding and shielded metal arc welding produce smoke containing particles of various types of oxides, which in some cases can lead to medical conditions like metal fume fever. The size of the particles in question tends to influence the toxicity of the fumes, with smaller particles presenting a greater danger. Additionally, many processes produce fumes and various gases, most commonly carbon dioxide, ozone and heavy metals, that can prove dangerous without proper ventilation and training. Furthermore, because the use of compressed gases and flames in many welding processes poses an explosion and fire risk, some common precautions include limiting the amount of oxygen in the air and keeping combustible materials away from the workplace. Welding fume extractors are often used to remove the fume from the source and filter the fumes through a HEPA filter.

Costs and trends

As an industrial process, the cost of welding plays a crucial role in manufacturing decisions. Many different variables affect the total cost, including equipment cost, labor cost, material cost, and energy cost. Depending on the process, equipment cost can vary, from inexpensive for methods like shielded metal arc welding and oxyfuel welding, to extremely expensive for methods like laser beam welding and electron beam welding. Because of their high cost, they are only used in high production operations. Similarly, because automation and robots increase equipment costs, they are only implemented when high production is necessary. Labor cost depends on the deposition rate (the rate of welding), the hourly wage, and the total operation time, including both time welding and handling the part. The cost of materials includes the cost of the base and filler material, and the cost of shielding gases. Finally, energy cost depends on arc time and welding power demand.

For manual welding methods, labor costs generally make up the vast majority of the total cost. As a result, many cost-savings measures are focused on minimizing the operation time. To do this, welding procedures with high deposition rates can be selected, and weld parameters can be fine-tuned to increase welding speed. Mechanization and automatization are often implemented to reduce labor costs, but this frequently increases the cost of equipment and creates additional setup time. Material costs tend to increase when special properties are necessary, and energy costs normally do not amount to more than several percent of the total welding cost.

In recent years, in order to minimize labor costs in high production manufacturing, industrial welding has become increasingly more automated, most notably with the use of robots in resistance spot welding (especially in the automotive industry) and in arc welding. In robot welding, mechanized devices both hold the material and perform the weld, and at first, spot welding was its most common application. But robotic arc welding has been increasing in popularity as technology has advanced. Other key areas of research and development include the welding of dissimilar materials (such as steel and aluminium, for example) and new welding processes, such as friction stir, magnetic pulse, conductive heat seam, and laser-hybrid welding. Furthermore, progress is desired in making more specialized methods like laser beam welding practical for more applications, such as in the aerospace and automotive industries. Researchers also hope to better understand the often unpredictable properties of welds, especially microstructure, residual stresses, and a weld's tendency to crack or deform.

Wood

2008/9 Schools Wikipedia Selection. Related subjects: Engineering; Organisms

Wood is hard, fibrous, lignified structural tissue produced as secondary xylem in the stems of woody plants, notably trees but also shrubs. It conducts water to the leaves and other growing tissues and acts as a support function, enabling plants to reach large sizes. Wood may also refer to other plant materials and tissues with comparable properties.

Wood is a heterogeneous, hygroscopic, cellular and anisotropic material. It is composed of fibers of cellulose (40% – 50%) and hemicellulose (15% – 25%) impregnated with lignin (15% – 30%).


Sections of tree trunk


A tree trunk as found at the Veluwe, The Netherlands


Artists can use wood to create delicate sculptures.

Wood has been used for millennia for many purposes. One of its primary uses is as fuel. It is also used as for making artworks, furniture, tools and weapons, and as a construction material.

Wood can be cut into straight planks and made into a hardwood floor (parquetry).

Wood has been an important construction material since humans began building shelters, houses and boats. Nearly all boats were made out of wood till the late 1800s, and wood remains in common use today in boat construction. New domestic housing in many parts of the world today is commonly made from timber-framed construction. In buildings made of other materials, wood will still be found as a supporting material, especially in roof construction, in interior doors and their frames, and as exterior cladding. Wood to be used for construction work is commonly known as *lumber* in North America. Elsewhere, *lumber* usually refers to felled trees, and the word for sawn planks ready for use is *timber*.

Wood unsuitable for construction in its native form may be broken down mechanically (into fibres or chips) or chemically (into cellulose) and used as a raw material for other building materials such as chipboard, engineered wood, hardboard, medium-density fiberboard (MDF), oriented strand board (OSB). Such wood derivatives are widely used: wood fibers are an important component of most paper, and cellulose is used as a component of some synthetic materials. Wood derivatives can also be used for kinds of flooring, for example laminate flooring.

Wood is also used for cutlery, such as chopsticks, toothpicks, and other utensils, like the wooden spoon.

Formation

A tree increases in diameter by the formation, between the old wood and the inner bark, of new woody layers which envelop the entire stem, living branches, and roots. Where there are clear seasons, this can happen in a discrete pattern, leading to what is known as growth rings, as can be seen on the end of a log. If these seasons are annual these growth rings are annual rings. Where there is no seasonal difference growth rings are likely to be indistinct or absent.

Within a growth ring it may be possible to see two parts. The part nearest the centre of the tree is more open textured and almost invariably lighter in colour than that near the outer portion of the ring. The inner portion is formed early in the season, when growth is comparatively rapid; it is known as early wood or spring wood. The outer portion is the late wood or summer wood, being produced in the summer. In white pines there is not much contrast in the different parts of the ring, and as a result the wood is very uniform in texture and is easy to work. In hard pines, on the other hand, the late wood is very dense and is deep-colored, presenting a very decided contrast to the soft, straw-colored early wood. In ring-porous woods each season's growth is always well defined, because the large pores of the spring abut on the denser tissue of the fall before. In the diffuse-porous woods, the demarcation between rings is not always so clear and in some cases is almost (if not entirely) invisible to the unaided eye.

Knots

A knot is a particular type of imperfection in a piece of timber, which reduces its strength, but which may be exploited for artistic effect. In a longitudinally-sawn plank, a knot will appear as a roughly circular "solid" (usually darker) piece of wood around which the roughly parallel fibres (grain) of the rest of the "flows" (parts and rejoins).


A knot on a tree at the Garden of the Gods public park in Colorado Springs, Colorado (October 2006).

A knot is actually a portion of a side branch (or a dormant bud) included in the wood of the stem or larger branch. The included portion is irregularly conical in shape (hence the roughly circular cross-section) with the tip at the point in stem diameter at which the plant's cambium was located when the branch formed as a bud. Within a knot, the fibre direction (grain) is up to 90 degrees different from the fibres of the stem, thus producing local cross grain.

During the development of a tree, the lower limbs often die, but may persist for a time, sometimes years. Subsequent layers of growth of the attaching stem are no longer intimately joined with the dead limb, but are grown around it. Hence, dead branches produce knots which are not attached, and likely to drop out after the tree has been sawn into boards.

In grading lumber and structural timber, knots are classified according to their form, size, soundness, and the firmness with which they are held in place. This firmness is affected by, among other factors, the length of time for which the branch was dead while the attaching stem continued to grow.

Knots materially affect cracking (known in the industry as checking) and warping, ease in working, and cleavability of timber. They are defects which weaken timber and lower its value for structural purposes where strength is an important consideration. The weakening effect is much more serious when timber is subjected to forces perpendicular to the grain and/or tension than where under load along the grain and/or compression. The extent to which knots affect the strength of a beam depends upon their position, size, number, direction of fibre, and condition. A knot on the upper side is compressed, while one on the lower side is subjected to tension. If there is a season check in the knot, as is often the case, it will offer little resistance to this tensile stress. Small knots, however, may be located along the neutral plane of a beam and increase the strength by preventing longitudinal shearing. Knots in a board or plank are least injurious when they extend through it at right angles to its broadest surface. Knots which occur near the ends of a beam do not weaken it. Sound knots which occur in the central portion one-fourth the height of the beam from either edge are not serious defects.

Knots do not necessarily influence the stiffness of structural timber. Only defects of the most serious character affect the elastic limit of beams. Stiffness and elastic strength are more dependent upon the quality of the wood fiber than upon defects in the beam. The effect of knots is to reduce the difference between the fibre stress at elastic limit and the modulus of rupture of beams. The breaking strength is very susceptible to defects. Sound knots do not weaken wood when subject to compression parallel to the grain.

For purposes for which appearance is more important than strength, such as wall panelling, knots are considered a benefit, as they add visual texture to the wood, giving it a more interesting appearance.

The traditional style of playing the Basque xylophon *txalaparta* involves hitting the right knots to obtain different tones.

Heartwood and sapwood

A section of a Yew branch showing 27 annual growth rings, pale sapwood and dark heartwood, and pith (centre dark spot). The dark radial lines are small knots.

Heartwood is wood that has died and become resistant to decay as a result of genetically programmed processes. It appears in a cross-section as a discolored circle, following annual rings in shape. Heartwood is usually much darker than living wood, and forms with age. Many woody plants do not form heartwood, but other processes, such as decay, can discolor wood in similar ways, leading to confusion. Some uncertainty still exists as to whether heartwood is truly dead, as it can still chemically react to decay organisms, but only once (Shigo 1986, 54).

Sapwood is living wood in the growing tree. All wood in a tree is first formed as sapwood. Its principal functions are to conduct water from the roots to the leaves and to store up and give back according to the season the food prepared in the leaves. The more leaves a tree bears and the more vigorous its growth, the larger the volume of sapwood required. Hence trees making rapid growth in the open have thicker sapwood for their size than trees of the same species growing in dense forests. Sometimes trees grown in the open may become of considerable size, 30 cm or more in diameter, before any heartwood begins to form, for example, in second-growth hickory, or open-grown pines.

The term *heartwood* derives solely from its position and not from any vital importance to the tree. This is evidenced by the fact that a tree can thrive with its heart completely decayed. Some species begin to form heartwood very early in life, so having only a thin layer of live sapwood, while in others the change comes slowly. Thin sapwood is characteristic of such trees as chestnut, black locust, mulberry, osage-orange, and sassafras, while in maple, ash, hickory, hackberry, beech, and pine, thick sapwood is the rule.

There is no definite relation between the annual rings of growth and the amount of sapwood. Within the same species the cross-sectional area of the sapwood is very roughly proportional to the size of the crown of the tree. If the rings are narrow, more of them are required than where they are wide. As the tree gets larger, the sapwood must necessarily become thinner or increase materially in volume. Sapwood is thicker in the upper portion of the trunk of a tree than near the base, because the age and the diameter of the upper sections are less.

When a tree is very young it is covered with limbs almost, if not entirely, to the ground, but as it grows older some or all of them will eventually die and are either broken off or fall off. Subsequent growth of wood may completely conceal the stubs which will however remain as knots. No matter how smooth and clear a log is on the outside, it is more or less knotty near the middle. Consequently the sapwood of an old tree, and particularly of a forest-grown tree, will be freer from knots than the heartwood. Since in most uses of wood, knots are defects that weaken the timber and interfere with its ease of working and other properties, it follows that sapwood, because of its position in the tree, may have certain advantages over heartwood.

It is remarkable that the inner heartwood of old trees remains as sound as it usually does, since in many cases it is hundreds of years, and in a few instances thousands of years, old. Every broken limb or root, or deep wound from fire, insects, or falling timber, may afford an entrance for decay, which, once started, may penetrate to all parts of the trunk. The larvae of many insects bore into the trees and their tunnels remain indefinitely as sources of weakness. Whatever advantages, however, that sapwood may have in this connection are due solely to its relative age and position.

If a tree grows all its life in the open and the conditions of soil and site remain unchanged, it will make its most rapid growth in youth, and gradually decline. The annual rings of growth are for many years quite wide, but later they become narrower and narrower. Since each succeeding ring is laid down on the outside of the wood previously formed, it follows that unless a tree materially increases its production of wood from year to year, the rings must necessarily become thinner as the trunk gets wider. As a tree reaches maturity its crown becomes more open and the annual wood production is lessened, thereby reducing still more the width of the growth rings. In the case of forest-grown trees so much depends upon the competition of the trees in their struggle for light and nourishment that periods of rapid and slow growth may alternate. Some trees, such as southern oaks, maintain the same width of ring for hundreds of years. Upon the whole, however, as a tree gets larger in diameter the width of the growth rings decreases.

There may be decided differences in the grain of heartwood and sapwood cut from a large tree, particularly one that is mature. In some trees, the wood laid on late in the life of a tree is softer, lighter, weaker, and more even-textured than that produced earlier, but in other species, the reverse applies. In a large log the sapwood, because of the time in the life of the tree when it was grown, may be inferior in hardness, strength, and toughness to equally sound heartwood from the same log.

Different woods

There is a strong relationship between the properties of wood and the properties of the particular tree that yielded it. For every tree species there is a range of density for the wood it yields. There is a rough correlation between density of a wood and its strength (mechanical properties). For example, while mahogany is a medium-dense hardwood which is excellent for fine furniture crafting, balsa is light, making it useful for model building. The densest wood may be black ironwood.

Wood is commonly classified as either softwood or hardwood. The wood from conifers (e.g. pine) is called softwood, and the wood from broad-leaved trees (e.g. oak) is called hardwood. These names are a bit misleading, as hardwoods are not necessarily hard, and softwoods are not necessarily soft. The well-known balsa (a hardwood) is actually softer than any commercial softwood. Conversely, some softwoods (e.g. yew) are harder than most hardwoods.

Wood products such as plywood are typically classified as engineered wood and not considered raw wood.

Colour (Colour)

In species which show a distinct difference between heartwood and sapwood the natural colour of heartwood is usually darker than that of the sapwood, and very frequently the contrast is conspicuous. This is produced by deposits in the heartwood of various materials resulting from the process of growth, increased possibly by oxidation and other chemical changes, which usually have little or no appreciable effect on the mechanical properties of the wood. Some experiments on very resinous Longleaf Pine specimens, however, indicate an increase in strength. This is due to the resin which increases the strength when dry. Such resin-saturated heartwood is called "fat lighter". Structures built of fat lighter are almost impervious to rot and termites; however they are very flammable. Stumps of old longleaf pines are often dug, split into small pieces and sold as kindling for fires. Stumps thus dug may actually remain a century or more since being cut. Spruce impregnated with crude resin and dried is also greatly increased in strength thereby.

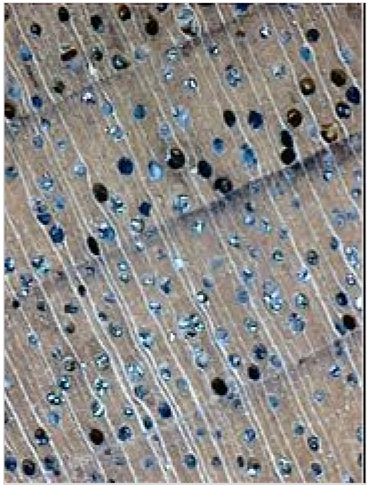

The wood of Coast Redwood is distinctively red in colour

Since the late wood of a growth ring is usually darker in colour than the early wood, this fact may be used in judging the density, and therefore the hardness and strength of the material. This is particularly the case with coniferous woods. In ring-porous woods the vessels of the early wood not infrequently appear on a finished surface as darker than the denser late wood, though on cross sections of heartwood the reverse is commonly true. Except in the manner just stated the colour of wood is no indication of strength.

Abnormal discolouration of wood often denotes a diseased condition, indicating unsoundness. The black check in western hemlock is the result of insect attacks. The reddish-brown streaks so common in hickory and certain other woods are mostly the result of injury by birds. The discolouration is merely an indication of an injury, and in all probability does not of itself affect the properties of the wood. Certain rot-producing fungi impart to wood characteristic colours which thus become symptomatic of weakness; however an attractive effect known as spalting produced by this process is often considered a desirable characteristic. Ordinary sap-staining is due to fungous growth, but does not necessarily produce a weakening effect.

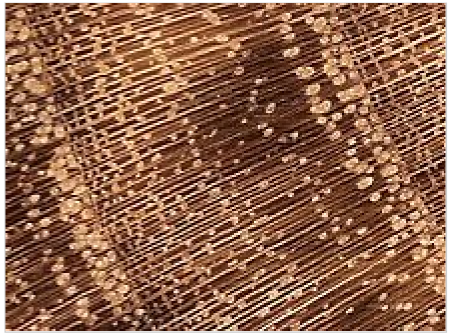
Structure

In coniferous or softwood species the wood cells are mostly of one kind, tracheids, and as a result the material is much more uniform in structure than that of most hardwoods. There are no vessels ("pores") in coniferous wood such as one sees so prominently in oak and ash, for example.


Magnified cross-section of a **diffuse-porous** hardwood wood (Black Walnut), showing the vessels, rays (white lines) and annual rings

The structure of the hardwoods is more complex. They are more or less filled with vessels: in some cases (oak, chestnut, ash) quite large and distinct, in others (buckeye, poplar, willow) too small to be seen plainly without a small hand lens. In discussing such woods it is customary to divide them into two large classes, *ring-porous* and *diffuse-porous*. In ring-porous species, such as ash, black locust, catalpa, chestnut, elm, hickory, mulberry, and oak, the larger vessels or pores (as cross sections of vessels are called) are localized in the part of the growth ring formed in spring, thus forming a region of more or less open and porous tissue. The rest of the ring, produced in summer, is made up of smaller vessels and a much greater proportion of wood fibres. These fibres are the elements which give strength and toughness to wood, while the vessels are a source of weakness.

In diffuse-porous woods the pores are scattered throughout the growth ring instead of being collected in a band or row. Examples of this kind of wood are basswood, birch, buckeye, maple, poplar, and willow. Some species, such as walnut and cherry, are on the border between the two classes, forming an intermediate group.


Black locust end grain, showing the **ring-porous** structure.

If a heavy piece of pine is compared with a light specimen it will be seen at once that the heavier one contains a larger proportion of late wood than the other, and is therefore considerably darker. The late wood of all species is denser than that formed early in the season, hence the greater the proportion of late wood the greater the density and strength. When examined under a microscope the cells of the late wood are seen to be very thick-walled and with very small cavities, while those formed first in the season have thin walls and large cavities. The strength is in the walls, not the cavities. In choosing a piece of pine where strength or stiffness is the important consideration, the principal thing to observe is the comparative amounts of early and late wood. The width of ring is not nearly so important as the proportion of the late wood in the ring.

It is not only the proportion of late wood, but also its quality, that counts. In specimens that show a very large proportion of late wood it may be noticeably more porous and weigh considerably less than the late wood in pieces that contain but little. One can judge comparative density, and therefore to some extent weight and strength, by visual inspection.


The twisty branch of a Lilac tree

No satisfactory explanation can as yet be given for the real causes underlying the formation of early and late wood. Several factors may be involved. In conifers, at least, rate of growth alone does not determine the proportion of the two portions of the ring, for in some cases the wood of slow growth is very hard and heavy, while in others the opposite is true. The quality of the site where the tree grows undoubtedly affects the character of the wood formed, though it is not possible to formulate a rule governing it. In general, however, it may be said that where strength or ease of working is essential, woods of moderate to slow growth should be chosen. But in choosing a particular specimen it is not the width of ring, but the proportion and character of the late wood which should govern.

In the case of the ring-porous hardwoods there seems to exist a pretty definite relation between the rate of growth of timber and its properties. This may be briefly summed up in the general statement that the more rapid the growth or the wider the rings of growth, the heavier, harder, stronger, and stiffer the wood. This, it must be remembered, applies only to ring-porous woods such as oak, ash, hickory, and others of the same group, and is, of course, subject to some exceptions and limitations.

In ring-porous woods of good growth it is usually the middle portion of the ring in which the thick-walled, strength-giving fibers are most abundant. As the breadth of ring diminishes, this middle portion is reduced so that very slow growth produces comparatively light, porous wood composed of thin-walled vessels and wood parenchyma. In good oak these large vessels of the early wood occupy from 6 to 10 per cent of the volume of the log, while in inferior material they may make up 25 per cent or more. The late wood of good oak, except for radial grayish patches of small pores, is dark colored and firm, and consists of thick-walled fibers which form one-half or more of the wood. In inferior oak, such fibre areas are much reduced both in quantity and quality. Such variation is very largely the result of rate of growth.

Wide-ringed wood is often called "second-growth", because the growth of the young timber in open stands after the old trees have been removed is more rapid than in trees in the forest, and in the manufacture of articles where strength is an important consideration such "second-growth" hardwood material is preferred. This is particularly the case in the choice of hickory for handles and spokes. Here not only strength, but toughness and resilience are important. The results of a series of tests on hickory by the U.S. Forest Service show that:

> "The work or shock-resisting ability is greatest in wide-ringed wood that has from 5 to 14 rings per inch (rings 1.8-5 mm thick), is fairly constant from 14 to 38 rings per inch (rings 0.7-1.8 mm thick), and decreases rapidly from 38 to 47 rings per inch (rings 0.5-0.7 mm thick). The strength at maximum load is not so great with the most rapid-growing wood; it is maximum with from 14 to 20 rings per inch (rings 1.3-1.8 mm thick), and again becomes less as the wood becomes more closely ringed. The natural deduction is that wood of first-class mechanical value shows from 5 to 20 rings per inch (rings 1.3-5 mm thick) and that slower growth yields poorer stock. Thus the inspector or buyer of hickory should discriminate against timber that has more than 20 rings per inch (rings less than 1.3 mm thick). Exceptions exist, however, in the case of normal growth upon dry situations, in which the slow-growing material may be strong and tough."

The effect of rate of growth on the qualities of chestnut wood is summarized by the same authority as follows:

> "When the rings are wide, the transition from spring wood to summer wood is gradual, while in the narrow rings the spring wood passes into summer

> wood abruptly. The width of the spring wood changes but little with the width of the annual ring, so that the narrowing or broadening of the annual ring is always at the expense of the summer wood. The narrow vessels of the summer wood make it richer in wood substance than the spring wood composed of wide vessels. Therefore, rapid-growing specimens with wide rings have more wood substance than slow-growing trees with narrow rings. Since the more the wood substance the greater the weight, and the greater the weight the stronger the wood, chestnuts with wide rings must have stronger wood than chestnuts with narrow rings. This agrees with the accepted view that sprouts (which always have wide rings) yield better and stronger wood than seedling chestnuts, which grow more slowly in diameter."

In diffuse-porous woods, as has been stated, the vessels or pores are scattered throughout the ring instead of collected in the early wood. The effect of rate of growth is, therefore, not the same as in the ring-porous woods, approaching more nearly the conditions in the conifers. In general it may be stated that such woods of medium growth afford stronger material than when very rapidly or very slowly grown. In many uses of wood, strength is not the main consideration. If ease of working is prized, wood should be chosen with regard to its uniformity of texture and straightness of grain, which will in most cases occur when there is little contrast between the late wood of one season's growth and the early wood of the next.

Monocot wood

Structural tissue resembling ordinary 'dicot' wood is produced by a number of monocot plants, and these are also usually called wood. Of these, the wood of the grass bamboo has considerable economic importance, larger culms being used in the manufacture of engineered flooring, panels and veneer. Other plant groups that produce woody tissue are palms, and members of the Liliales, such as Dracaena and Cordyline. With all these woods, the structure and composition of the structural tissue is quite different from ordinary wood.

Water content

Water occurs in living wood in three conditions, namely: (1) in the cell walls, (2) in the protoplasmic contents of the cells, and (3) as free water in the cell cavities and spaces. In heartwood it occurs only in the first and last forms. Wood that is thoroughly air-dried retains from 8-16% of water in the cell walls, and none, or practically none, in the other forms. Even oven-dried wood retains a small percentage of moisture, but for all except chemical purposes, may be considered absolutely dry.

The general effect of the water content upon the wood substance is to render it softer and more pliable. A similar effect of common observation is in the softening action of water on paper or cloth. Within certain limits the greater the water content the greater its softening effect.

Drying produces a decided increase in the strength of wood, particularly in small specimens. An extreme example is the case of a completely dry spruce block 5 cm in section, which will sustain a permanent load four times as great as that which a green block of the same size will support.

The greatest increase due to drying is in the ultimate crushing strength, and strength at elastic limit in endwise compression; these are followed by the modulus of rupture, and stress at elastic limit in cross-bending, while the modulus of elasticity is least affected.

Wood as fuel

Wood is burned as a fuel mostly in rural areas of the world. Hard wood is preferred over softwood because it creates less smoke and burns longer. Adding a woodstove or fireplace to a home adds ambiance and warmth.