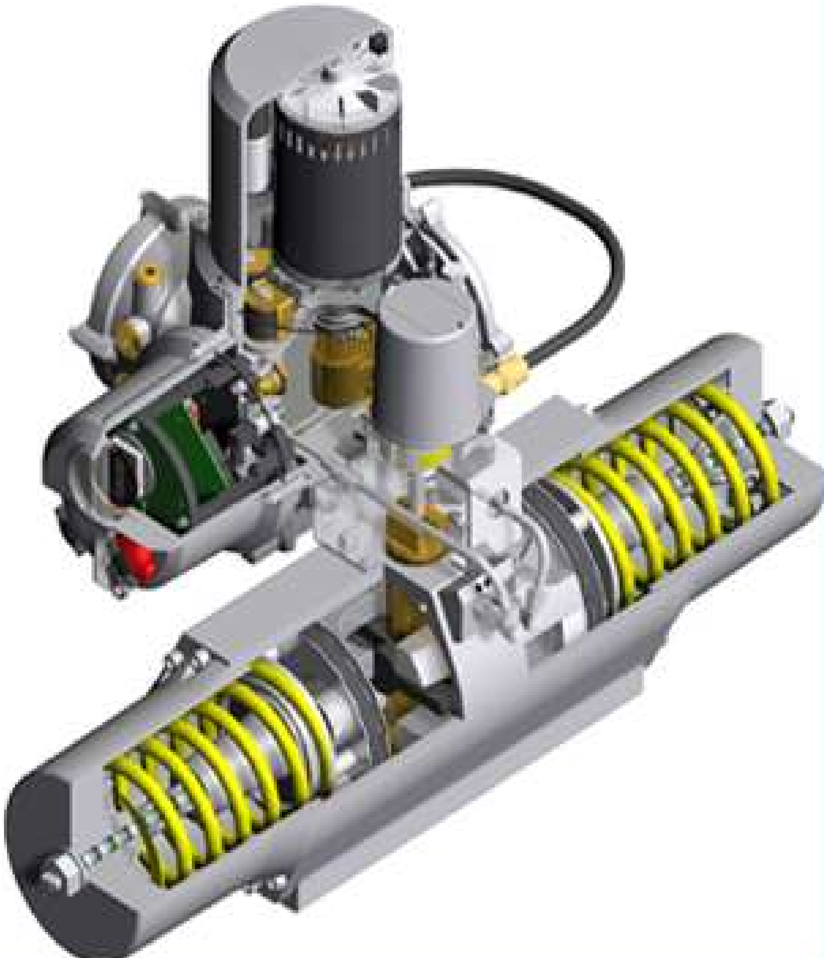

Handbook of

# Detectors, Sensors and Actuator Mechanical Devices

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## Chapter- 1

# Metal Detector

A U.S. Army soldier uses a metal detector to search for weapons and ammunition in Iraq

A **metal detector** is a device which responds to metal that may not be readily apparent.

The simplest form of a metal detector consists of an oscillator producing an alternating current that passes through a coil producing an alternating magnetic field. If a piece of electrically conductive metal is close to the coil, eddy currents will be induced in the metal, and this produces an alternating magnetic field of its own. If another coil is used to measure the magnetic field (acting as a magnetometer), the change in the magnetic field due to the metallic object can be detected.

The first industrial metal detectors were developed in the 1960s and were used extensively for mining and other industrial applications. Uses include de-mining (the detection of land mines), the detection of weapons such as knives and guns, especially in airport security, geophysical prospecting, archaeology and treasure hunting. Metal detectors are also used to detect foreign bodies in food, and in the construction industry to detect steel reinforcing bars in concrete and pipes and wires buried in walls and floors.

# History and development

Toward the end of the 19th century, many scientists and engineers used their growing knowledge of electrical theory in an attempt to devise a machine which would pinpoint metal. The use of such a device to find ore-bearing rocks would give a huge advantage to any miner who employed it. The German physicist Heinrich Wilhelm Dove invented the induction balance system, which was incorporated into metal detectors a hundred years later. Early machines were crude, used a lot of battery power, and worked only to a very limited degree. Alexander Graham Bell used such a device to attempt to locate a bullet lodged in the chest of American President James Garfield in 1881; the attempt was unsuccessful because the metal coil spring bed Garfield was lying on confused the detector.

## Modern developments

The modern development of the metal detector began in the 1930s. Gerhard Fisher had developed a system of radio direction-finding, which was to be used for accurate navigation. The system worked extremely well, but Fisher noticed that there were anomalies in areas where the terrain contained ore-bearing rocks. He reasoned that if a radio beam could be distorted by metal, then it should be possible to design a machine which would detect metal using a search coil resonating at a radio frequency. In 1937 he applied for, and was granted, the first patent for a metal detector. However, it was one Lieutenant Jozef Stanislaw Kosacki, a Polish officer attached to a unit stationed in St Andrews, Fife, Scotland, during the early years of World War II, who refined the design into a practical Polish mine detector. They were heavy, ran on vacuum tubes, and needed separate battery packs.

The design invented by Kosacki was used extensively during the clearance of the German mine fields during the Second Battle of El Alamein when 500 units were shipped to Field Marshal Montgomery to clear the minefields of the retreating Germans, and later used during the Allied invasion of Sicily, the Allied invasion of Italy and the Invasion of Normandy. As it was a wartime research operation to create and refine the design of the detector, the knowledge that Kosacki created the first practical metal detector was kept secret for over 50 years.

After the war, there were plenty of surplus mine detectors on the market; they were bought up by relic hunters who used them for fun and profit. This helped to form metal detecting into a hobby.

### Further refinements

Many manufacturers of these new devices brought their own ideas to the market. Whites Electronics of Oregon began in the '50s by building a machine called the Oremaster Geiger Counter. Another leader in detector technology was Charles Garrett, who pioneered the BFO (Beat Frequency Oscillator) machine. With the invention and development of the transistor in the '50s and '60s, metal detector manufacturers and designers made smaller lighter machines with improved circuitry, running on small battery packs. Companies sprang up all over the USA and Britain to supply the growing demand.

Modern top models are fully computerized, using integrated circuit technology to allow the user to set sensitivity, discrimination, track speed, threshold volume, notch filters, etc., and hold these parameters in memory for future use. Compared to just a decade ago, detectors are lighter, deeper-seeking, use less battery power, and discriminate better.

Larger portable metal detectors are used by archaeologists and treasure hunters to locate metallic items, such as jewelry, coins, bullets, and other various artifacts buried shallowly underground.

### Discriminators

The biggest technical change in detectors was the development of the induction-balance system. This system involved two coils that were electrically balanced. When metal was introduced to their vicinity, they would become unbalanced. What allowed detectors to discriminate between metals was the fact that every metal has a different phase response when exposed to alternating current. Scientists had long known of this fact by the time detectors were developed that could selectively detect desirable metals, while ignoring undesirable ones.

Even with discriminators, it was still a challenge to avoid undesirable metals, because some of them have similar phase responses e.g. tinfoil and gold, particularly in alloy form. Thus, improperly tuning out certain metals increased the risk of passing over a valuable find. Another disadvantage of discriminators was that they reduced the sensitivity of the machines.

### New coil designs

Coil designers also tried out innovative designs. The original induction balance coil system consisted of two identical coils placed on top of one another. Compass Electronics produced a new design: two coils in a D shape, mounted back-to-back to form a circle. This system was widely used in the 1970s, and both concentric and D type (or widescan as they became known) had their fans. Another development was the invention of detectors which could cancel out the effect of mineralization in the ground. This gave greater depth, but was a non-discriminate mode. It worked best at lower frequencies than those used before, and frequencies of 3 to 20 kHz were found to produce the best results. Many detectors in the 1970s had a switch which enabled the user to switch between the discriminate mode and the non-discriminate mode. Later developments switched electronically between both modes. The development of the induction balance detector would ultimately result in the motion detector, which constantly checked and balanced the background mineralization.

## Pulse induction

A pulse induction metal detector with an array of coils

At the same time, developers were looking at using a different technique in metal detection called Pulse Induction. Unlike the Beat Frequency Oscillator or the Induction Balance machines which both used a uniform alternating current at a low frequency, the pulse induction machine simply fired a high-voltage pulse of signal into the ground. In the absence of metal, the 'spike' decayed at a uniform rate, and the time it took to fall to zero volts could be accurately measured. However, if metal was present when the machine fired, a small current would flow in the metal, and the time for the voltage to drop to zero would be increased. These time differences were minute, but the improvement in electronics made it possible to measure them accurately and identify the presence of metal at a reasonable distance. These new machines had one major advantage: they were completely impervious to the effects of mineralization, and rings and other jewelry could now be located even under highly-mineralized black sand.

# Uses

## Archaeology

In England and Wales metal detecting is legal provided that permission is granted by the landowner, and that the area is not a Scheduled Ancient Monument, a site of special scientific interest (SSSI), or covered by elements of the Countryside Stewardship Scheme. Items discovered which fall within the definition of treasure must be reported to the coroner or a place designated by the coroner for treasure. The voluntary reporting of finds which do not qualify as treasure to the Portable Antiquities Scheme or the UK Detector Finds Database is encouraged.

The situation in Scotland is very different. Under the Scots law principle of *bona vacantia*, the Crown has claim over any object of any material value where the original owner cannot be traced. There is also no 300 year limit to Scottish finds. Any artifact found, whether by metal detector survey or from an archaeological excavation, must be reported to the Crown through the Treasure Trove Advisory Panel at the National Museums of Scotland. The panel then determines what will happen to the artifacts. Reporting is not voluntary, and failure to report the discovery of historic artifacts is a criminal offence in Scotland.

**As a hobby**


This 156 ounce nugget was found by an individual prospector in the Southern California Desert using a metal detector

There are six major types of hobbyist activities involving metal detectors:

- Coin shooting is looking for coins after an event involving many people, like a baseball game, or simply looking for any old coins. Serious coin shooters will spend hours, days and months doing historical research to locate long lost sites that have the potential to give up historical and collectible coins.
- Prospecting is looking for valuable metals like gold and silver in their natural forms, such as nuggets or flakes.
- General metal detecting is very similar to coin shooting except that the metal detectorist is after any type of historical artifact. Metal detectorists may be dedicated to preserving historical artifacts, and often have considerable expertise. Coins, bullets, buttons, axe heads, and buckles are just a few of the items that are commonly found by relic hunters; in general the potential is far greater in Europe and Asia than many other parts of the world. More valuable finds in Britain alone

include the Staffordshire Hoard of Anglo-Saxon gold, sold for £3,285,000, the gold Celtic Newark Torc, the Ringlemere Cup, West Bagborough Hoard, Milton Keynes Hoard, Roman Crosby Garrett Helmet, Stirling Hoard, Collette Hoard and thousands of smaller finds.

- Beach combing is hunting for lost coins or jewelry on a beach. Beach hunting can be as simple or as complicated as one wishes to make it. Many dedicated beach hunters also familiarize themselves with tide movements and beach erosion. There are two main techniques for beach hunting. The first one is called "gridding", which is when you search in a pattern. For example, you start from the beach line, and work your way down to the shoreline, move to the side a little, and repeat the process. The next technique is called "Random searching". Random searching is when you walk around the beach in no particular pattern, hoping to cover more ground.
- Metal detecting clubs across the United States, United Kingdom and Canada exist for hobbyists to learn from others, show off finds from their hunts and to learn more about the hobby.

## Security screening

Metal detectors at an airport

A series of aircraft hijackings led the Finnish company Outokumpu to adapt mining metal detectors, still housed in a large cylindrical pipe, to the purpose of screening airline passengers as they walked through. The development of these systems continued in a spin off company and systems branded as Metor Metal Detectors evolved in the form of the rectangular gantry now standard in airports. In common with the developments in other uses of metal detectors both alternating current and pulse systems are used, and the design of the coils and the electronics has moved forward to improve the discrimination of these systems. In 1995 systems such as the Metor 200 appeared with the ability to indicate the approximate height of the metal object above the ground, enabling security personnel to more rapidly locate the source of the signal. Smaller hand held metal detectors are also used to locate a metal object on a person more precisely.

## Industrial metal detectors

Industrial metal detectors are used in the pharmaceutical, food, beverage, textile, garment, plastics, chemicals, lumber, and packaging industries.

Contamination of food by metal shards from broken processing machinery during the manufacturng process is a major safety issue in the food industry. Metal detectors for this purpose are widely used and integrated into the production line.

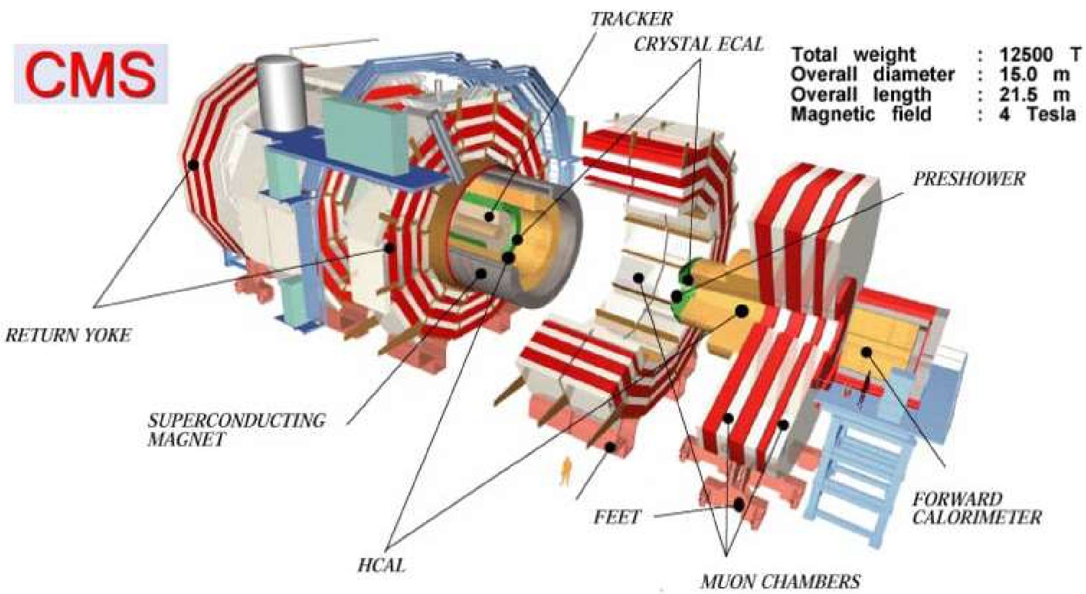
Current practice at garment or apparel industry plants is to apply metal detecting after the garments are completely sewn and before garments are packed to check whether there is any metal contamination (needle, broken needle, etc) in the garments. This needs to be done for safety reasons.

## Civil engineering

In civil engineering special metal detectors (cover meters) are used to locate rebars. Rebar detectors are less sophisticated devices that can only locate metallic objects below the surface.

## Chapter- 2

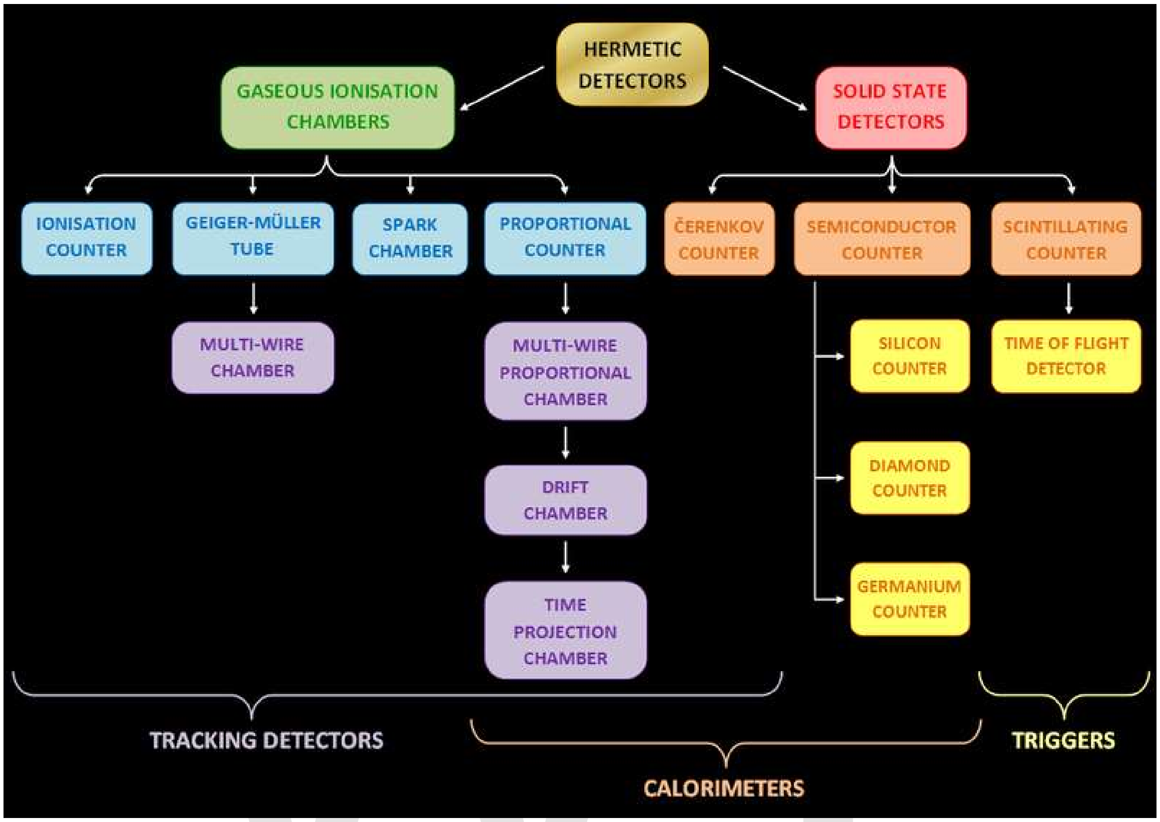
# Particle Detector




The Compact Muon Solenoid (CMS) is an example of a large particle detector. Notice the person for scale.

In experimental and applied particle physics, nuclear physics, and nuclear engineering, a **particle detector**, also known as a **radiation detector**, is a device used to detect, track, and/or identify high-energy particles, such as those produced by nuclear decay, cosmic radiation, or reactions in a particle accelerator. Modern detectors are also used as calorimeters to measure the energy of the detected radiation. They may also be used to measure other attributes such as momentum, spin, charge etc. of the particles.

# Description




Summary of Particle Detectors

Detectors designed for modern accelerators are huge, both in size and in cost. The term *counter* is often used instead of *detector*, when the detector counts the particles but does not resolve its energy or ionization. Particle detectors usually can also track ionizing radiation (high energy photons or even visible light). If their main purpose is radiation measurement, they are called *radiation detectors*, but as photons are also (massless) particles, the term *particle detector* is still correct.

## Examples and types

Many of the detectors invented and used so far are ionization detectors (of which gaseous ionization detectors and semiconductor detectors are most typical) and scintillation detectors; but other, completely different principles have also been applied, like Čerenkov light and transition radiation.

Cloud chamber with visible tracks from ionizing radiation (short, thick: α-particles; long, thin: β-particles

**Historical Examples**

- Bubble chamber
- Wilson cloud chamber (diffusion chamber)
- Photographic plates

**Detectors for Radiation Protection**

- Dosimeter
- Electroscope (miniature electroscopes are used as portable dosimeters)

**Commonly used detectors for Particle and Nuclear Physics**

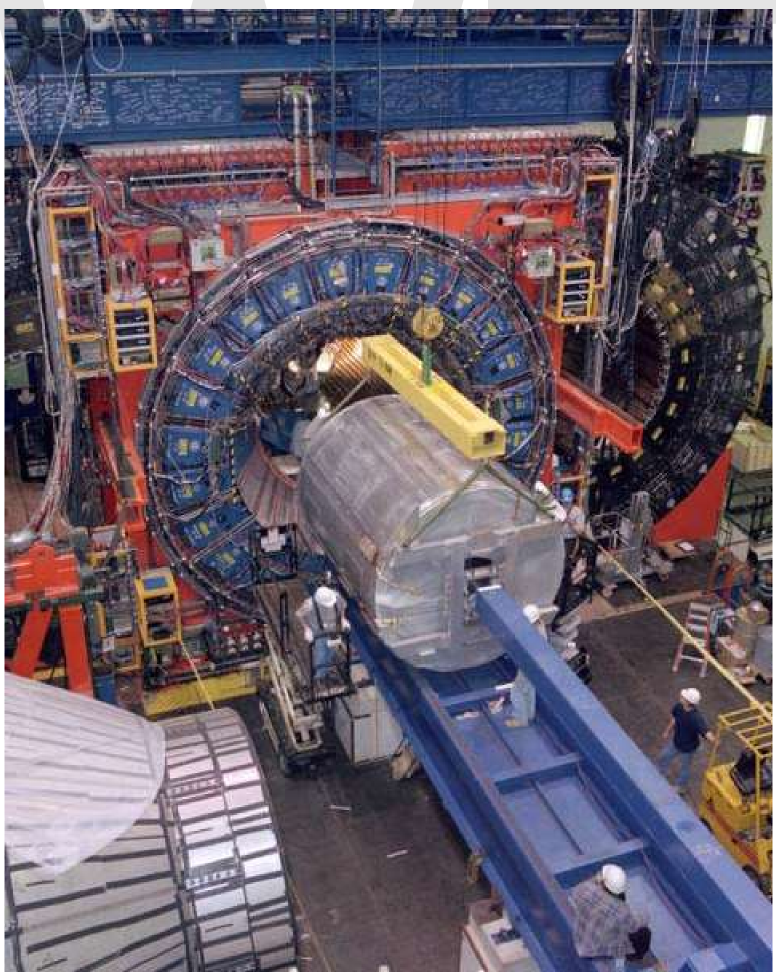
- Gaseous ionization detectors
    - Ionization chamber
    - Proportional counter
        - Multiwire Proportional Chamber
        - Drift chamber
        - Time projection chamber

- Geiger-Müller tube
- Spark chamber

- Solid-state detectors
  - Cherenkov detector
    - RICH (Ring Imaging Cherenkov Detector)
  - Scintillation counter and associated Photomultiplier or Photodiode / Avalanche photodiode
    - Lucas cell
    - Time of flight detector
  - Semiconductor detector
    - Silicon Vertex Detector
  - Transition radiation detector

- Calorimeters

## Hermetic Detectors

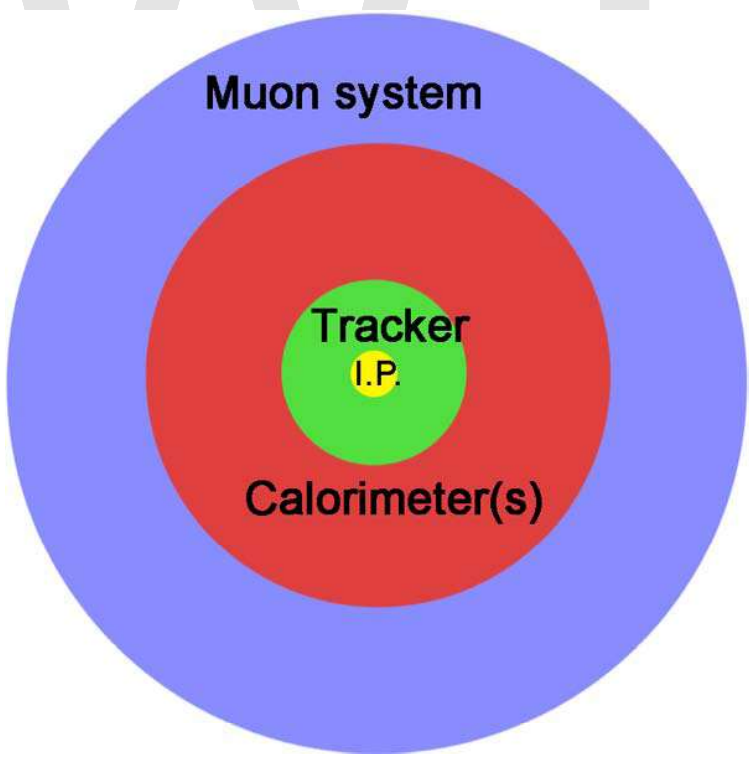


A currently-operating hermetic detector, the Collider Detector at Fermilab

In particle physics, a **hermetic detector** (also called a **4π detector**) is a particle detector designed to observe all possible decay products of an interaction between subatomic particles in a collider by covering as large an area around the interaction point as possible and incorporating multiple types of sub-detectors. They are typically roughly cyllindrical, with different types of detectors wrapped around each other; each detector type specializes in particular particles so that almost any particle will be detected and identified. Such detectors are called "hermetic" because they are designed to let as few particles as possible escape; the name "4π detector" comes from the fact that such detectors are designed to cover nearly all of the 4π steradians of solid angle around the interaction point.

The first such detector was the Mark I at the Stanford Linear Accelerator Center, and the basic design has been used for all subsequent collider detectors. Prior to the building of the Mark I, it was thought that most particle decay products would have relatively low transverse momentum (i.e. momentum perpendicular to the beamline), so that detectors could cover this area only. However, it was learned at the Mark I and subsequent experiments that most fundamental particle interactions at colliders involve very large exchanges of energy and therefore large transverse momenta are not uncommon; for this reason, large angular coverage is critical for modern particle physics.

## Components




A schematic of the basic components of a hermetic detector; I.P. refers to the region containing the interaction point for the colliding particles. This is a cross section of the typical cyllindrical design.

There are three main components of a hermetic detector. From the inside out, the first is a **tracker**, which measures the momentum of charged particles as they curve in a magnetic field. Next there are one or more **calorimeters**, which measure the energy of most charged and neutral particles by absorbing them in dense material, and a **muon system** which measures the one type of particle that is not stopped through the calorimeters and can still be detected. Each component may have several different specialized sub-components.

### Trackers

The tracking system plots the helix traced by a charged particle that curves in a magnetic field by localizing it in space in finely-segmented layers of detecting material, usually silicon. The degree to which the particle curves is inversely proportional to its momentum perpendicular to the beam, while the degree to which it drifts in the direction of the beam axis gives its momentum in that direction.

### Calorimeters

Calorimeters slow particles down and absorb their energy into a material, allowing that energy to be measured. They are often divided into two types: the electromagnetic calorimeter that specializes in absorbing particles that interact electromagnetically, and the hadronic calorimeter that can detect hadrons, which interact via the strong nuclear force. A hadronic detector is required in particular to detect heavy neutral particles.

### Muon system

Of all the known stable particles, only muons and neutrinos pass through the calorimeter without losing most or all of their energy. Neutrinos are practically undetectable, and their existence must be inferred, but muons (which are charged) can be measured by an additional tracking system outside the calorimeters.

## Particle identification

Most particles have unique combinations of signals left in each detector sub-system, allowing different particles to be identified. For example, an electron is charged and interacts electromagnetically, so it is tracked by the tracker and then deposits all of its energy in the (electromagnetic) calorimeter. By contrast, a photon is neutral and interacts electromagnetically, so it deposits its energy in the calorimeter without leaving a track.

## Installations of particle detectors

### At colliders

- At CERN
    - for the LHC

        - CMS
        - ATLAS
        - ALICE
        - LHCb
    - for the LEP
        - Aleph
        - Delphi
        - L3
        - Opal
    - for the SPS
        - The COMPASS Experiment
        - Gargamelle
        - NA49
- At Fermilab
    - for the Tevatron
        - CDF
        - D0
- At DESY
    - for HERA
        - H1
        - HERA-B
        - HERMES
        - ZEUS
- At BNL
    - for the RHIC
        - PHENIX
        - Phobos (physics)
        - STAR
- At SLAC
    - for the PeP-II
        - BaBar
    - for the SLC
        - SLD
- At Cornell
    - for CESR
        - CLEO
        - CUSB
- Others
    - MECO from UC Irvine

### Under construction

- For ILC
    - CALICE

**Without colliders**

- Super-Kamiokande
- AMANDA
- CMDS

## Chapter- 3

# Smoke Detector

A **smoke detector** is a device that detects smoke, typically as an indicator of fire. Commercial, industrial, and mass residential devices issue a signal to a fire alarm system, while household detectors, known as smoke alarms, generally issue a local audible and/or visual alarm from the detector itself.

Smoke detectors are typically housed in a disk-shaped plastic enclosure about 150 millimetres (6 in) in diameter and 25 millimetres (1 in) thick, but the shape can vary by manufacturer or product line. Most smoke detectors work either by optical detection (photoelectric) or by physical process (ionization), while others use both detection methods to increase sensitivity to smoke. Sensitive alarms can be used to detect, and thus deter, smoking in areas where it is banned such as toilets and schools. Smoke detectors in large commercial, industrial, and residential buildings are usually powered by a central fire alarm system, which is powered by the building power with a battery backup. However, in many single family detached and smaller multiple family housings, a smoke alarm is often powered only by a single disposable battery.

## History

The first automatic electric fire alarm was invented in 1890 by Francis Robbins Upton (U.S. patent no. 436,961). Upton was an associate of Thomas Edison, but there is no evidence that Edison contributed to this project.

In the late 1930s the Swiss physicist Walter Jaeger tried to invent a sensor for poison gas. He expected that gas entering the sensor would bind to ionized air molecules and thereby alter an electric current in a circuit in the instrument. His device failed: small concentrations of gas had no effect on the sensor's conductivity. Frustrated, Jaeger lit a cigarette—and was soon surprised to notice that a meter on the instrument had registered a drop in current. Smoke particles had apparently done what poison gas could not. Jaeger's experiment was one of the advances that paved the way for the modern smoke detector.

It was 30 years, however, before progress in nuclear chemistry and solid-state electronics made a cheap sensor possible. While home smoke detectors were available during most of the 1960s, the price of these devices was rather high. Before that, alarms were so expensive that only major businesses and theaters could afford them.

The first truly affordable home smoke detector was invented by Duane D. Pearsall in 1965, featuring an individual battery powered unit that could be easily installed and replaced. The first units for mass production came from Duane Pearsall's company, Statitrol Corporation, in Lakewood, Colorado.

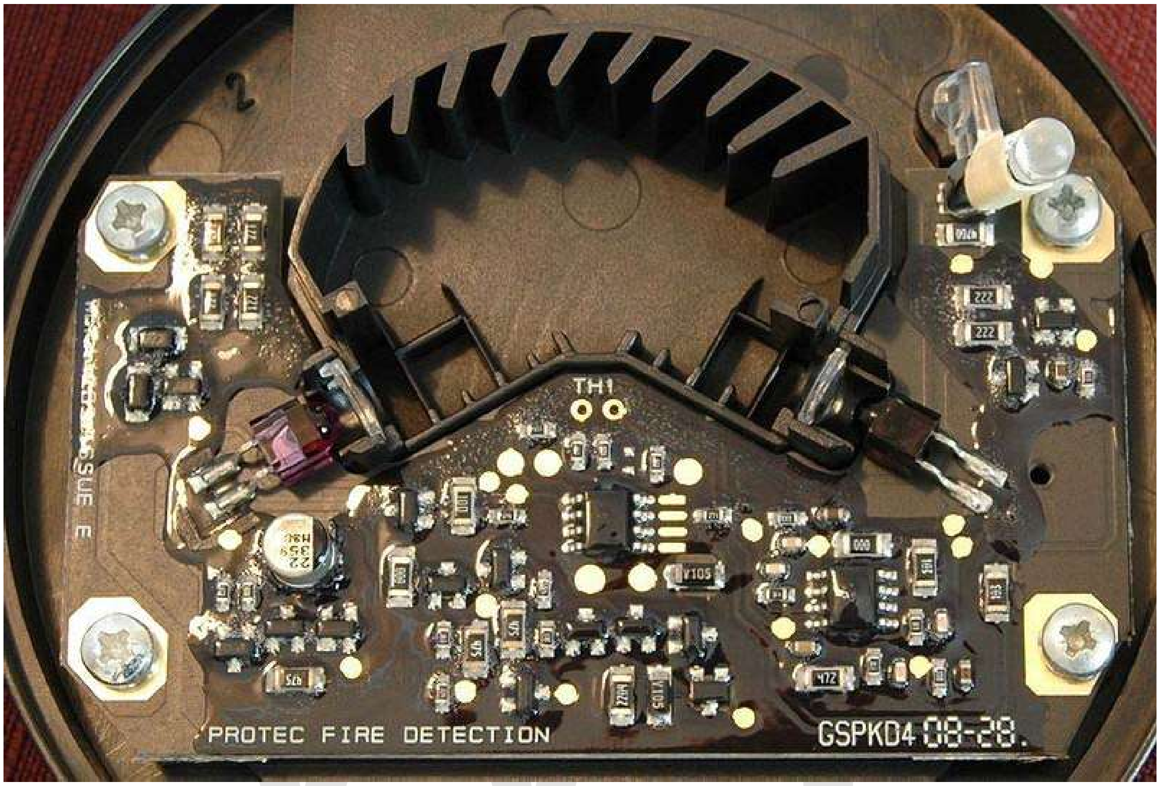
These first units were made from strong fire resistant steel and shaped much like a bee's hive. The battery was a rechargeable specialized unit created by Gates Energy. The need for a quick replace battery didn't take long to show itself and the rechargeable was replaced with a pair of AA batteries along with a plastic shell encasing the detector. The small assembly line sent close to 500 units per day before Statitrol sold its invention to Emerson Electric in 1980 and Sears's retailers picked up full distribution of the 'now required in every home' smoke detector.

The first commercial smoke detectors came to market in 1969. Today they are installed in 93% of U.S. homes and 85% of UK homes. However it is estimated that any given time over 30% of these alarms don't work, as users remove the batteries, or forget to replace them.

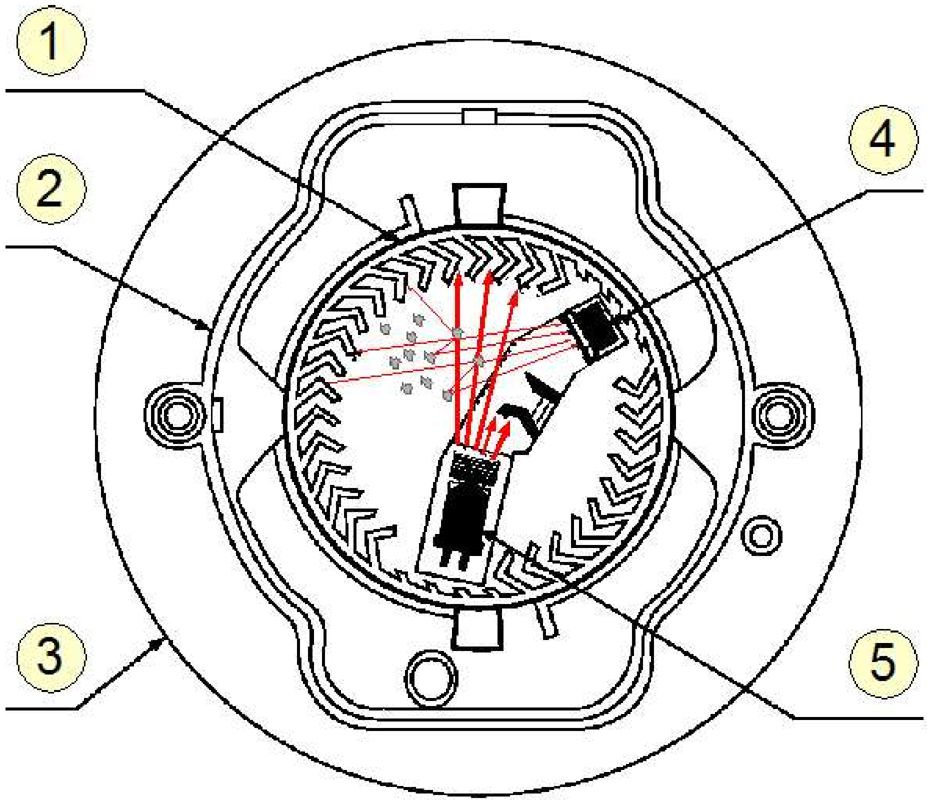
Although commonly attributed to NASA, smoke detectors were not invented as a result of the space program, though a variant with adjustable sensitivity was developed for Skylab.

# Design

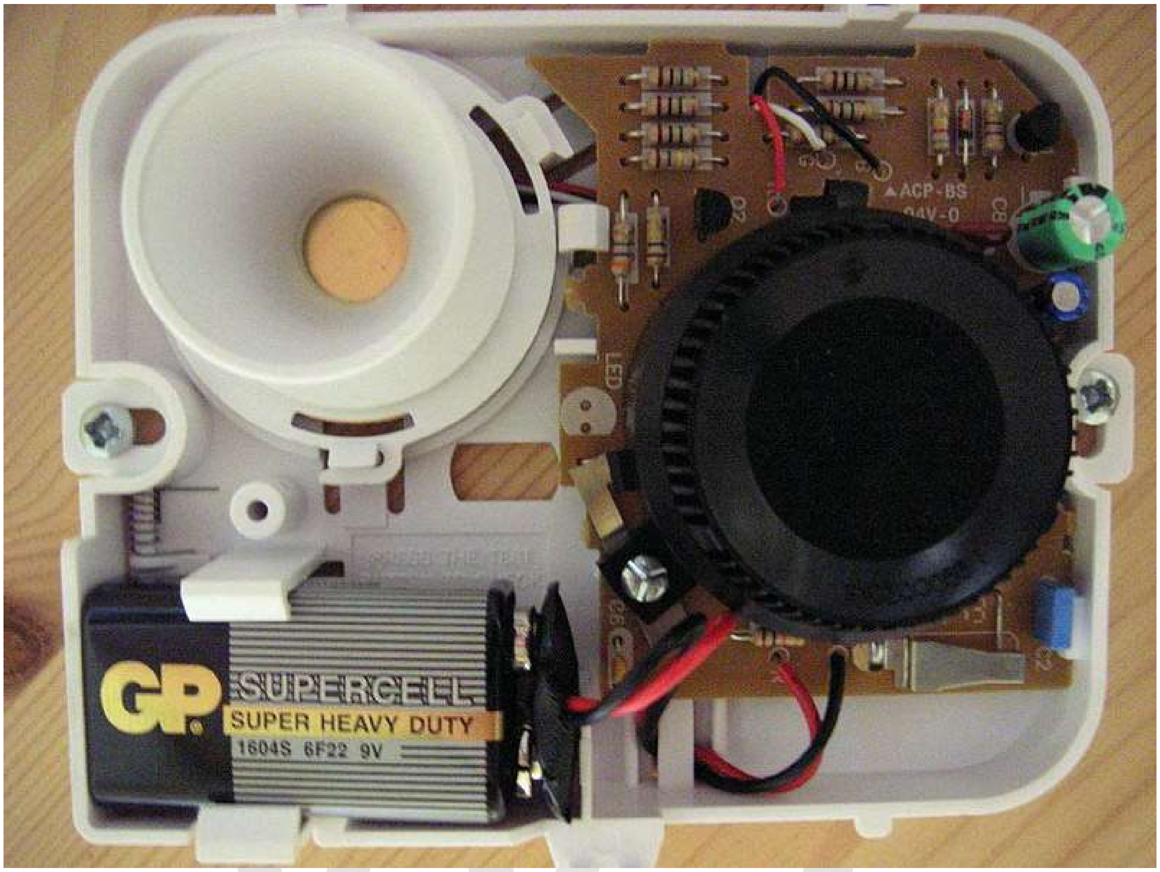
## Optical




Optical Smoke Detector with the cover removed




Optical Smoke Detector
1: Optical chamber
2: Cover
3: Case moulding
4: Photodiode (detector)
5: Infrared LED




Inside a basic ionization smoke detector. The black, round structure at the right is the ionization chamber. The white, round structure at the upper left is the piezoelectric buzzer that produces the alarm sound.

An optical detector is a light sensor. When used as a smoke detector, it includes a light source (incandescent bulb or infrared LED), a lens to collimate the light into a beam, and a photodiode or other photoelectric sensor at an angle to the beam as a light detector. In the absence of smoke, the light passes in front of the detector in a straight line. When smoke enters the optical chamber across the path of the light beam, some light is scattered by the smoke particles, directing it at the sensor and thus triggering the alarm.

Also seen in large rooms, such as a gymnasium or an auditorium, are devices to detect a projected beam. A unit on the wall sends out a beam, which is either received by a receiver or reflected back via a mirror. When the beam is less visible to the "eye" of the sensor, it sends an alarm signal to the fire alarm control panel.

Optical smoke detectors are quick in detecting particulate (smoke) generated by smoldering (cool, smoky) fires. Many independent tests indicate that optical smoke detectors typically detect particulates (smoke) from hot, flaming fires approximately 30 seconds later than ionization smoke alarms.

They are less sensitive to false alarms from steam or cooking fumes generated in kitchen or steam from the bathroom than are ionization smoke alarms. For the aforementioned reason, they are often referred to as 'toast proof' smoke alarms.

## Ionization

This type of detector is cheaper than the optical detector; however, it is sometimes rejected because it is more prone to false (nuisance) alarms than photoelectric smoke detectors. It can detect particles of smoke that are too small to be visible. It includes about 37 kBq or 1 μCi of radioactive americium-241 ($^{241}Am$), corresponding to about 0.3 μg of the isotope. The radiation passes through an ionization chamber, an air-filled space between two electrodes, and permits a small, constant current between the electrodes. Any smoke that enters the chamber absorbs the alpha particles, which reduces the ionization and interrupts this current, setting off the alarm.

$^{241}Am$, an alpha emitter, has a half-life of 432 years. This means that it does not have to be replaced during the useful life of the detector, and also makes it safe for people at home, since it is only slightly radioactive. Alpha radiation, as opposed to beta and gamma, is used for two additional reasons: Alpha particles have high ionization, so sufficient air particles will be ionized for the current to exist, and they have low penetrative power, meaning they will be stopped by the plastic of the smoke detector and/or the air. About one percent of the emitted radioactive energy of $^{241}Am$ is gamma radiation.

## Air-sampling

An air-sampling smoke detector is capable of detecting microscopic particles of smoke. Most air-sampling detectors are aspirating smoke detectors, which work by actively drawing air through a network of small-bore pipes laid out above or below a ceiling in parallel runs covering a protected area. Small holes drilled into each pipe form a matrix of holes (sampling points), providing an even distribution across the pipe network. Air samples are drawn past a sensitive optical device, often a solid-state laser, tuned to detect the extremely small particles of combustion. Air-sampling detectors may be used to trigger an automatic fire response, such as a gaseous fire suppression system, in high-value or mission-critical areas, such as archives or computer server rooms.

Most air-sampling smoke detection systems are capable of a higher sensitivity than spot type smoke detectors and provide multiple levels of alarm threshold, such as Alert, Action, Fire 1 and Fire 2. Thresholds may be set at levels across a wide range of smoke levels. This provides earlier notification of a developing fire than spot type smoke detection, allowing manual intervention or activation of automatic suppression systems before a fire has developed beyond the smoldering stage, thereby increasing the time available for evacuation and minimizing fire damage.

## Carbon monoxide and carbon dioxide detection

Some smoke alarms use a carbon dioxide sensor or carbon monoxide sensor in order to detect extremely dangerous products of combustion. However, not all smoke detectors that are advertised with such gas sensors are actually able to warn of poisonous levels of those gases in the absence of a fire.

## Performance differences

Optical or "toast-proof" smoke detectors are generally quicker in detecting particulate (smoke) generated by smoldering (cool, smokey) fires. Ionization smoke detectors are generally quicker in detecting particulate (smoke) generated by flaming (hot) fires.

According to fire tests conformant to EN 54, normally the $CO_2$ cloud from smoke can be detected before particulate.

Obscuration is a unit of measurement that has become the standard definition of smoke detector [Sensitivity (electronics) [sensitivity]]. Obscuration is the effect that smoke has on reducing visibility. Higher concentrations of smoke result in higher obscuration levels, lowering visibility.

**Typical smoke detector obscuration ratings**

| Type of Detector | Obscuration Level |
|---|---|
| Ionization | 2.6–5.0% obs/m (0.8–1.5% obs/ft) |
| Photoelectric | 6.5–13.0% obs/m (2–4% obs/ft) |
| Beam | 3% obs/m (0.9% obs/ft) |
| Aspirating | 0.005–20.5% obs/m (0.0015–6.25% obs/ft) |
| Laser | 0.06–6.41% obs/m (0.02–2.0% obs/ft) |

## Commercial smoke detectors

An integrated locking mechanism for commercial building doors. Inside an enclosure are a locking device, smoke detector and power supply.

Commercial smoke detectors are either conventional or analog addressable, and are wired up to security monitoring systems or fire alarm control panels (FACP). These are the most common type of detector, and usually cost a lot more than a household smoke alarms. They exist in most commercial and industrial facilities, such as high rises, ships and trains. These detectors don't need to have built in alarms, as alarm systems can be controlled by the connected FACP, which will set off relevant alarms, and can also implement complex functions such as a staged evacuation.

### Conventional

The word Conventional is slang used in to distinguish the method used to communicate with the control unit from that used by addressable detectors whose methods were unconventional at the time of their introduction. So called “Conventional Detectors” cannot be individually identified by the control unit and resemble an electrical switch in their information capacity. These detectors are connected in parallel to the signaling path or (initiating device circuit) so that the current flow is monitored to indicate a closure of the circuit path by any connected detector when smoke or other similar environmental stimulus sufficiently influences any detector. The resulting increase in current flow is interpreted and processed by the control unit as a confirmation of the presence of smoke and a fire alarm signal is generated.

### Addressable

This type of installation gives each detector on a system an individual number, or address. Thus, addressable detectors allow an FACP, and therefore fire fighters, to know the exact location of an alarm where the address is indicated on a diagram.

Analog addressable detectors provide information about the amount of smoke in their detection area, so that the FACP can decide itself, if there is an alarm condition in that area (possibly considering day/night time and the readings of surrounding areas). These are usually more expensive than autonomous deciding detectors.

## Standalone smoke alarms

The main function of a standalone smoke alarm is to alert persons at risk. Several methods are used and documented in industry specifications published by Underwriters Laboratories Alerting methods include:

- Audible tones
    - Usually around 3200 Hz due to component constraints
    - 85 dBA at 10 feet
- Spoken voice alert
- Visual strobe lights
    - 110 candela output
- Tactile stimulation, e.g., bed or pillow shaker (No standards exist as of 2008 for tactile stimulation alarm devices.)

Some models have a hush or temporary silence feature that allows silencing without removing the battery. This is especially useful in locations where false alarms can be relatively common (i.e. due to "toast burning") or users could remove the battery permanently to avoid the annoyance of false alarms, but removing the battery permanently is strongly discouraged.

While current technology is very effective at detecting smoke and fire conditions, the deaf and hard of hearing community has raised concerns about the effectiveness of the alerting function in awakening sleeping individuals in certain high risk groups such as the elderly, those with hearing loss and those who are intoxicated. Between 2005 and 2007, research sponsored by the United States' National Fire Protection Association (NFPA) has focused on understanding the cause of a higher number of deaths seen in such high risk groups. Initial research into the effectiveness of the various alerting methods is sparse. Research findings suggest that a low frequency (520 Hz) square wave output is significantly more effective at awakening high risk individuals. Wireless Wi-Safe smoke and carbon monoxide detectors linked to alert mechanisms such as vibrating pillow pads, strobes and remote warning handsets have been found to support the groups above.

## Batteries


Photoelectric smoke detector equipped with strobe light for the hearing impaired

Most residential smoke detectors run on 9-volt alkaline or carbon-zinc batteries. When these batteries run down, the smoke detector becomes inactive. Most smoke detectors will signal a low-battery condition. The alarm may chirp at intervals if the battery is low, though if there is more than one unit within earshot, it can be hard to locate. It is common, however, for houses to have smoke detectors with dead batteries. It is estimated, in the UK, that over 30% of smoke alarms may have dead or removed batteries. As a result, public information campaigns have been created to remind people to change smoke detector batteries regularly. In Australia, for example, it is advertised that all smoke alarm batteries should be replaced on the first day of April every year. In regions using daylight saving time, these campaigns may suggest that people change their batteries when they change their clocks or on a birthday.

Some detectors are also being sold with a lithium battery that can run for about 7 to 10 years, though this might actually make it less likely for people to change batteries, since their replacement is needed so infrequently. By that time, the whole detector may need to be replaced. Though relatively expensive, user-replaceable 9-volt lithium batteries are also available.

Common NiMH and NiCd rechargeable batteries have a high self-discharge rate, making them unsuitable for use in smoke detectors. This is true even though they may provide much more power than alkaline batteries if used soon after charging, such as in a portable stereo. Also, a problem with rechargeable batteries is a rapid voltage drop at the end of their useful charge. This is of concern in devices such as smoke detectors, since the battery may transition from "charged" to "dead" so quickly that the low-battery warning period from the detector is either so brief as to go unnoticed, or may not occur at all.

The NFPA, recommends that home-owners replace smoke detector batteries with a new battery at least once per year, when it starts chirping (a signal that its charge is low), or when it fails a test, which the NFPA recommends to be carried out at least once per month by pressing the "test" button on the alarm.

## Reliability

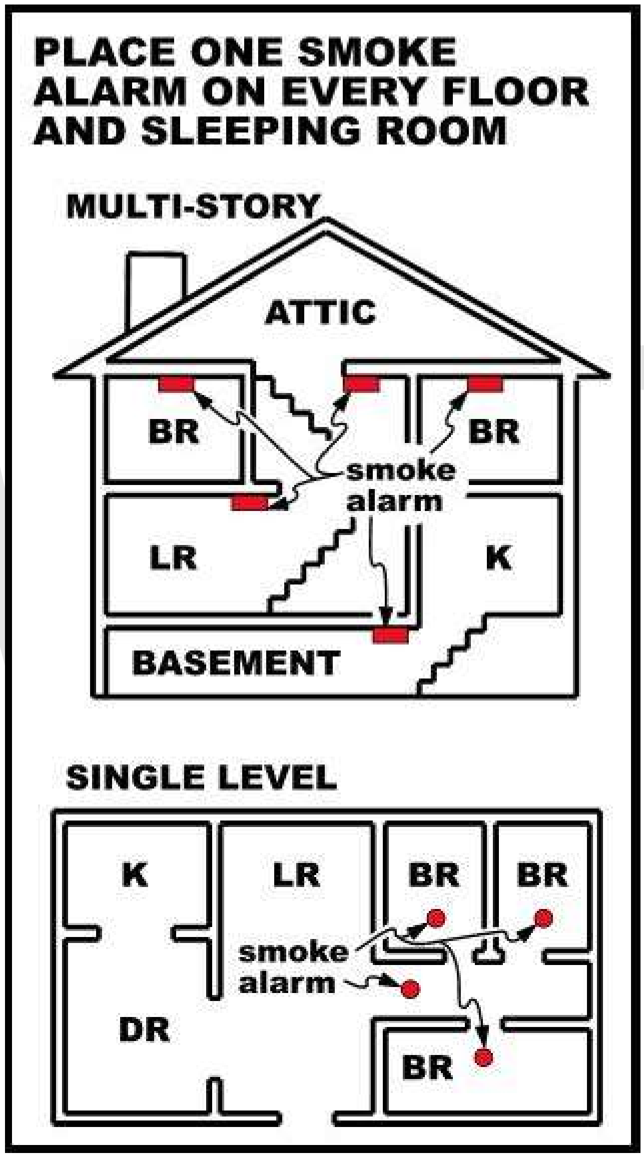
In 2004, NIST issued a comprehensive report that concludes, among other things, that "smoke alarms of either the ionization type or the photoelectric type consistently provided time for occupants to escape from most residential fires", and "consistent with prior findings, ionization type alarms provided somewhat better response to flaming fires than photoelectric alarms, and photoelectric alarms provided (often) considerably faster response to smoldering fires than ionization type alarms".

The NFPA strongly recommends the replacement of home smoke alarms every 10 years. Smoke alarms become less reliable with time, primarily due to aging of their electronic components, making them susceptible to nuisance false alarms. In ionization type alarms, decay of the $^{241}$Am radioactive source is a negligible factor, as its half-life is far greater than the expected useful life of the alarm unit.

Regular cleaning can prevent false alarms caused by the build up of dust or other objects such as flies, particularly on optical type alarms as they are more susceptible to these factors. A vacuum cleaner can be used to clean ionization and optical detectors externally and internally. However, on commercial ionisation detectors it is not recommended for a lay person to clean internally. To reduce false alarms caused by cooking fumes, use an optical or 'toast proof' alarm near the kitchen.

A jury in the United States District Court for the Northern District of New York decided in 2006 that First Alert and its parent company, BRK Brands, was liable for millions of dollars in damages because the ionization technology in the smoke alarm in the Hackert's house was defective, failing to detect the slow-burning fire and choking smoke that filled the home as the family slept.

## Installation and placement




A 2007 U.S. guide to placing smoke detectors, suggesting that one be placed on every floor of a building, and in each bedroom.

In the United States, most state and local laws regarding the required number and placement of smoke detectors are based upon standards established in Article 72 of the NFPA fire code.

Laws governing the installation of smoke detectors vary depending on the locality. Homeowners with questions or concerns regarding smoke detector placement may contact their local fire marshal or building inspector for assistance. However, some rules and guidelines for existing homes are relatively consistent throughout the developed world. For example, Canada and Australia require a building to have a working smoke detector on every level. The United States requires smoke detectors on every habitable level and within the vicinity of all bedrooms. Habitable levels include attics that are tall enough to allow access.

In new construction, minimum requirements are typically more stringent. All smoke detectors must be hooked directly to the electrical wiring, be interconnected and have a battery backup. In addition, smoke detectors are required either inside or outside every bedroom, depending on local codes. Smoke detectors on the outside will detect fires more quickly, assuming the fire does not begin in the bedroom, but the sound of the alarm will be reduced and may not wake some people. Some areas also require smoke detectors in stairways, main hallways and garages.

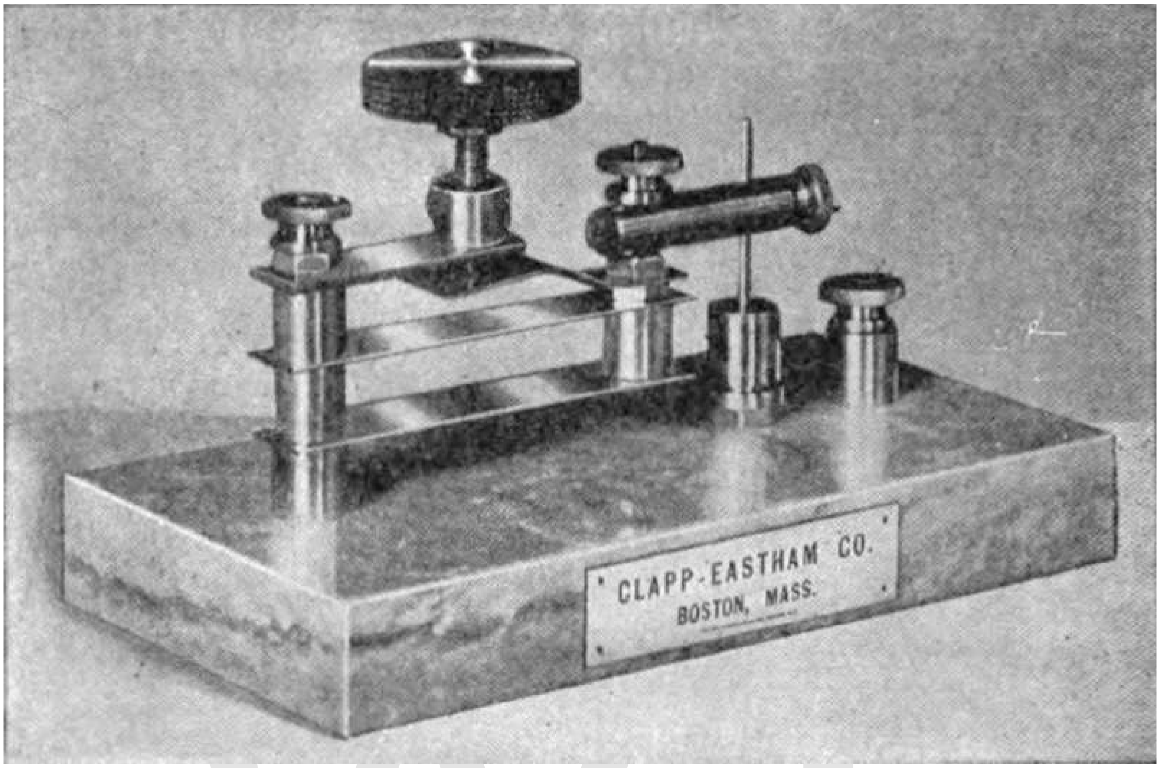
Wired units with a third "interconnect" wire allow a dozen or more detectors to be connected, so that if one detects smoke, the alarms will sound on all the detectors in the network, improving the chances that occupants will be alerted, even if they are behind closed doors or if the alarm is triggered one or two floors from their location. Wired interconnection may only be practical for use in new construction, especially if the wire needs to be routed in areas that are inaccessible without cutting open walls and ceilings. As of the mid-2000s, development has begun on wirelessly networking smoke alarms, using technologies such as ZigBee, which will allow interconnected alarms to be easily retrofitted in a building without costly wire installations. Some wireless systems using Wi-Safe technology will also detect smoke or carbon monoxide through the detectors, which simultaneously alarm themselves with vibrating pads, strobes and remote warning handsets. As these systems are wireless they can easily be transferred from one property to another.

In the UK the placement of detectors are similar however the installation of smoke alarms in new builds need to comply to the British Standards BS5839 pt6. BS 5839: Pt.6: 2004 recommends that a new-build property consisting of no more than 3 floors (less than 200sqm per floor)) should be fitted with a Grade D, LD2 system. Building Regulations in England, Wales and Scotland recommend that BS 5839: Pt.6 should be followed, but as a minimum a Grade D, LD3 system should be installed. Building Regulations in Northern Ireland require a Grade D, LD2 system to be installed, with smoke alarms fitted in the escape routes and the main living room and a heat alarm in the kitchen, this standard also requires all detectors to have a main supply and a battery back up.

## Chapter- 4

# Cat’s-whisker Detector

Galena cat's whisker detector




Precision cat's whisker detector using iron pyrite crystal, early 1900s. The crystal is inside the metal capsule under the vertical needle "whisker" *(right)*.

A **cat's whisker detector** (also called a **crystal detector**, a term that now also refers to modern semiconductor diodes) is an antique electronic component consisting of a thin wire that lightly touches a crystal of semiconducting mineral to make a crude contact-junction rectifier. Developed by early radio researchers Jagadish Chandra Bose, G. W. Pickard and others, this device was used as the detector in early crystal radios, from about 1906 through World War 2. It gave this type of radio receiver its name. It was the first type of semiconductor diode, and in fact the first semiconductor electronic device. The term cat's whisker was also sometimes used to describe the crystal receiver itself. Cat's whisker detectors are obsolete and are now only used in antique or antique-reproduction radios.

## Description

A small portable crystal radio, with a cat's whisker detector visible at top

The wire touching the surface of the crystal formed a crude and unstable metal–semiconductor point-contact P-N junction, forming a Schottky barrier diode. This junction conducts electric current in only one direction and resists current flowing in the other direction. In a crystal radio its function was to rectify the radio signal, converting it from alternating current to a pulsing direct current, to extract the audio signal (modulation) from the radio frequency carrier wave.

Only certain sites on the crystal surface functioned as rectifying junctions. Thus, the device was very sensitive to the exact geometry and pressure of contact between wire and crystal. Therefore it was made adjustable, and a usable point of contact was found by trial and error before each use. The wire was suspended from a moveable arm, and was dragged across the crystal face by the operator until the device began functioning. In a crystal radio, the operator would tune the radio to a strong local station, and then adjust the cat's whisker until the station was heard in the radio's earphones. This required some skill and a great deal of patience; even then a good contact could easily be lost by the slightest vibration. An alternate method of adjustment was to use a battery-operated buzzer to generate a test signal. The spark at the buzzer's contacts functioned as a weak radio transmitter, so when the crystal began functioning the buzz could be heard in the earphones, and the buzzer was turned off. The temperamental, unreliable action of the crystal detector was a barrier to its acceptance as a standard component in commercial radio equipment, and was one reason for its rapid replacement by vacuum tubes after 1920. Frederick Seitz, a later semiconductor researcher, wrote:

Such variability, bordering on what seemed the mystical, plagued the early history of crystal detectors and caused many of the vacuum tube experts of a later generation to regard the art of crystal rectification as being close to disreputable.

## Crystal

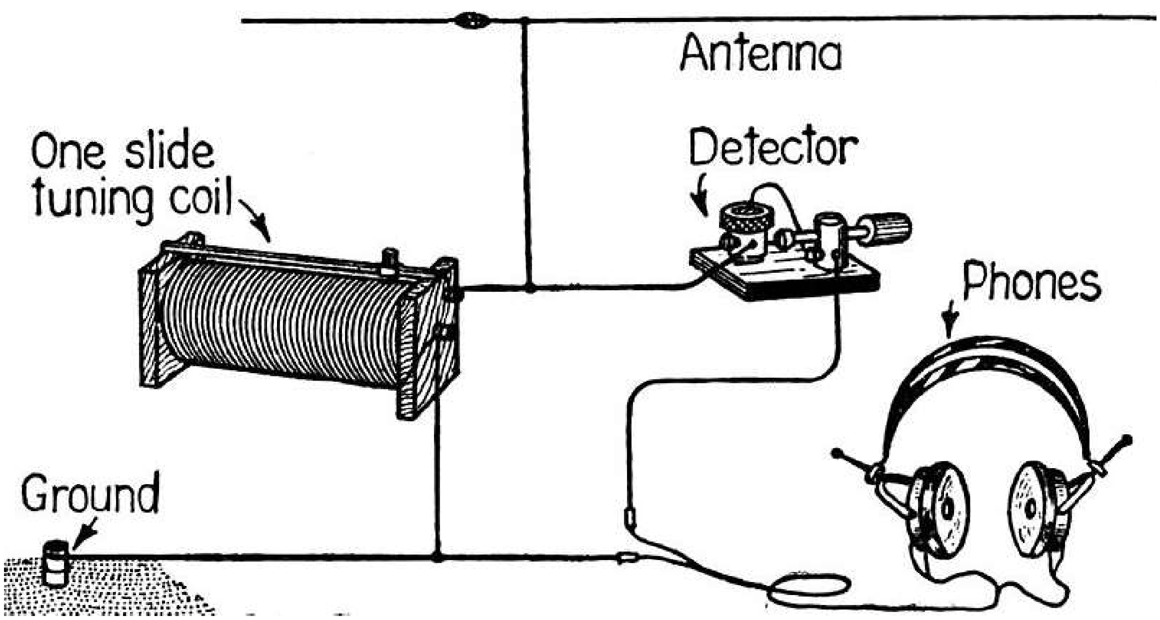
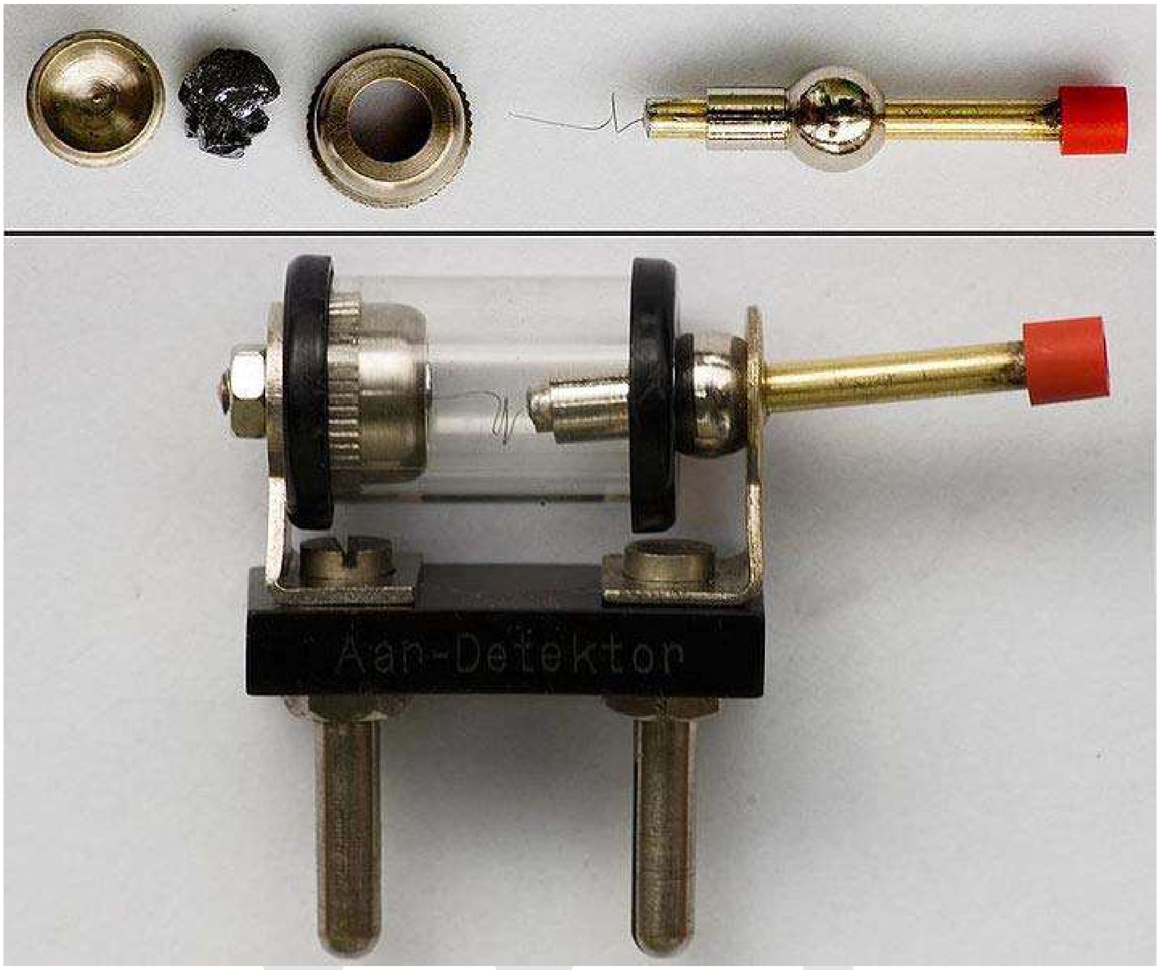



Diagram of a crystal set using a cat's-whisker detector

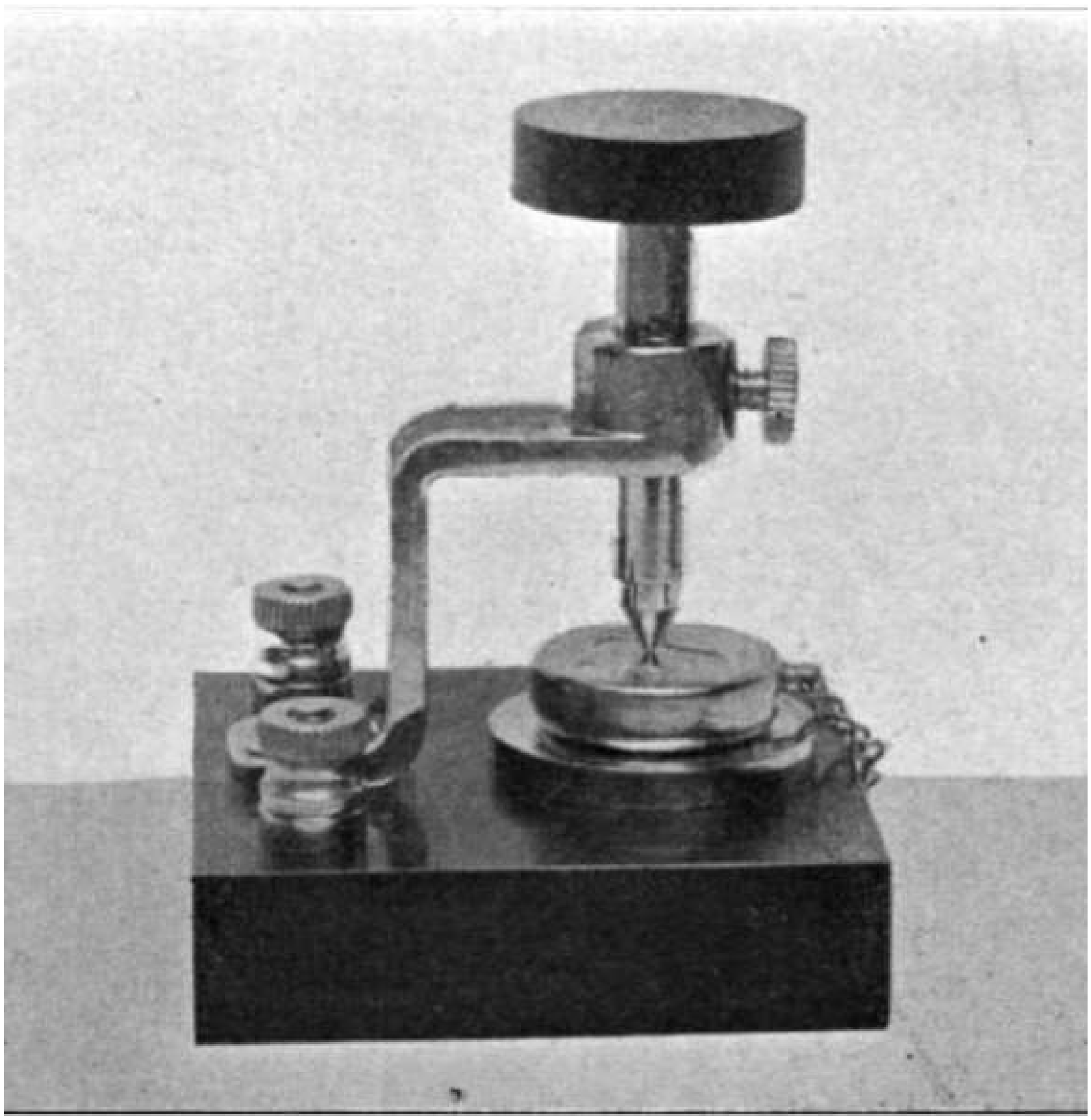



Modern galena cat's whisker detector, showing parts. The galena crystal *(upper left)* is held in the metal capsule with a screw cap, leaving its face exposed.

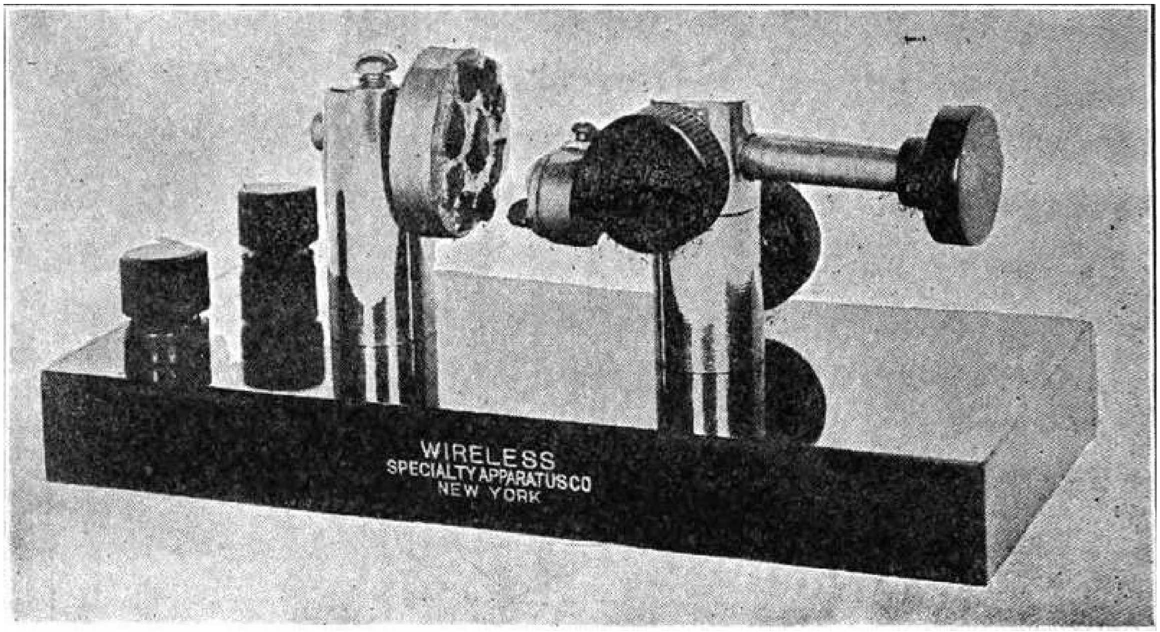
A natural mineral crystal forms the semiconductor side of the junction. The most common crystal used was galena (PbS, lead sulfide), a naturally occurring ore of lead. Galena is a semiconductor with a small bandgap of about 0.4 eV, and is used without treatment directly as it is mined. Galena with good detecting properties is rare and has no reliable visual characteristics distinguishing it from galena samples with poor detecting properties. A rough pebble of detecting mineral about the size of a pea was mounted in a metal cup, which formed one side of the circuit. The electrical contact between the cup and the crystal had to be good, because this contact must *not* act as a second rectifying junction. To make good contact with the crystal, it was either clamped with setscrews or mounted in solder. Because the relatively high temperature of tin-lead solder can damage many crystals, a low melting point (well under 200°F) alloy such as Wood's metal was used. One surface was left exposed to allow contact with the cat's whisker wire.

### Whisker

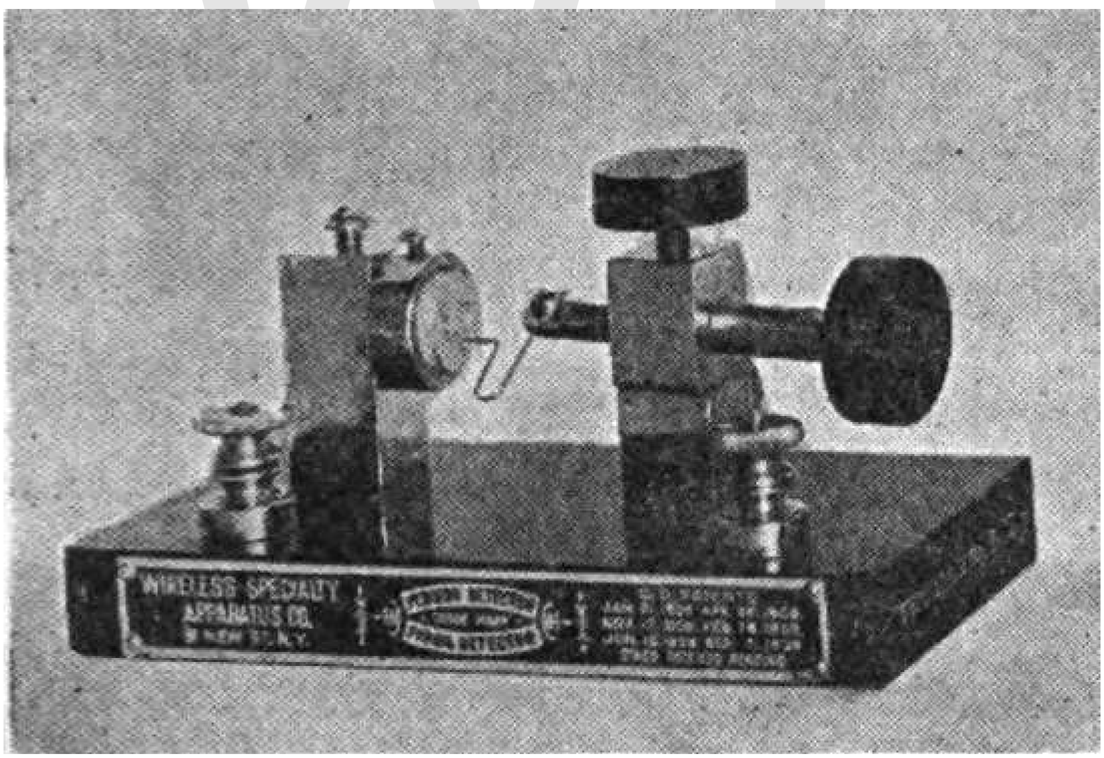
The "cat's whisker", a springy piece of thin metal wire, formed the metal side of the junction. Phosphor bronze wire of about 30 gauge was commonly used because it had the right amount of springiness. It was mounted on an adjustable arm with an insulated handle so that the entire exposed surface of the crystal could be probed from many directions to try to find the most sensitive spot. Cat's whiskers in simple detectors were straight or curved, but most cat's whiskers had a coiled section that acted as a spring near the arm, with a straight section just in back of the tip that probes the crystal. The crystal required just the right gentle pressure by the wire; too much pressure caused the device to conduct in both directions. Precision detectors often used a metal needle instead of a cat's whisker, mounted on a thumbscrew-operated leaf spring to adjust the pressure applied.



The first crystal detector commercially produced, Pickard's silicon detector, from 1906. Silicon didn't require the delicate "cat's whisker" contact.

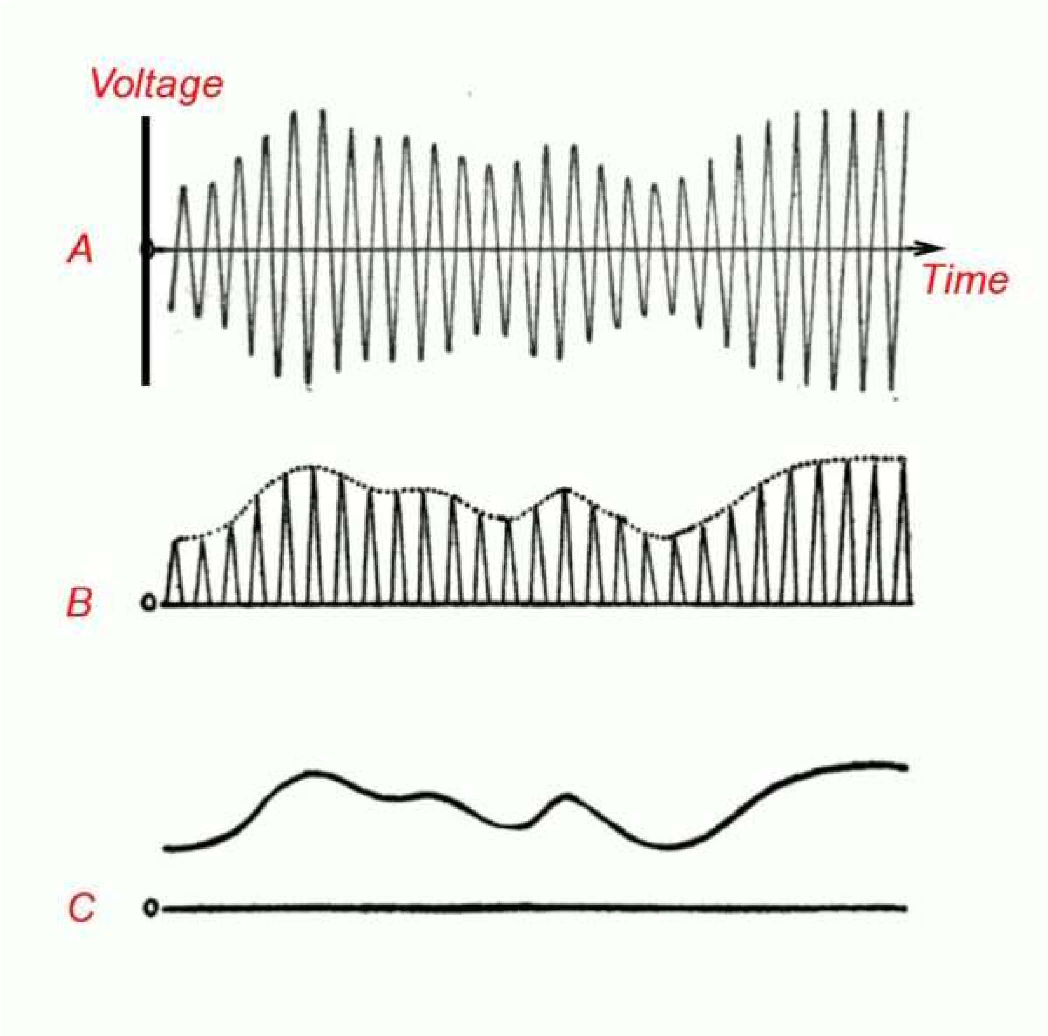



A "Perikon" detector, 1914. Instead of a wire-to-crystal contact, this had a crystal-to-crystal contact. The bornite crystal *(right)* on the adjustable arm was moved forward until it touched one of the zincite crystals on the carousel *(left)*. Multiple zincite crystals were provided because the zincite was vulnerable to damage from atmospheric electricity.



Pyrite detector

## Types



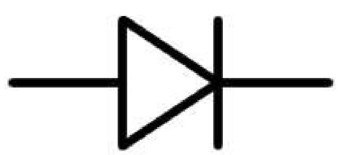

How a crystal detector works in a radio receiver. *(A)* The amplitude modulated radio signal from the receiver's tuning section. The rapid oscillations are the radio frequency carrier wave. The audio signal (the sound) is contained in the slow variations (modulation) of the size of the waves. This signal cannot be converted to sound by the earphone, because the audio excursions are the same on both sides of the axis, averaging out to zero, resulting in no net motion of the earphone's diaphragm. *(B)* The crystal conducts current in only one direction, stripping off the oscillations on one side of the signal, leaving a pulsing direct current whose amplitude does not average zero but varies with the audio signal. *(C)* A bypass capacitor across the earphone smoothes the waveform, removing the radio frequency carrier pulses, leaving the audio signal.

Historically, many other minerals and compounds besides galena were used for the crystal, the most important being iron pyrite ("fool's gold", iron disulfide), silicon, molybdenite ($MoS_2$), and silicon carbide (carborundum, SiC). Some were used with gold or graphite "cat's whiskers". Another type had a crystal-to-crystal junction instead of a "cat's whisker", with two crystals mounted facing each other. One crystal was moved forward on an adjustable mount until the crystal faces touched. The most common of these was a zincite-bornite ($ZnO$-$Cu_5FeS_4$) junction trade-named *Perikon*, but zincite-chalcopyrite, silicon-arsenic and silicon-antimony junctions were also used. The goal of researchers was to find junctions that were not as sensitive to vibration and unreliable as galena and pyrite. Some of these other junctions, particularly carborundum, were stable enough that they used a more permanent spring-loaded contact rather than a "cat's whisker". For this reason carborundum detectors were preferred in large commercial wireless stations, and military and shipboard stations which were subject to vibration from waves and gunnery exercises. Another quality desired was the ability to withstand high currents without damage, because in wireless stations the fragile detector junction could be "burned out" by atmospheric electric charge from the antenna, or high radio frequency current leaking into the receiver from the powerful spark-gap transmitter during transmissions. Carborundum detectors, which used large area contacts, were also particularly robust in this regard.

To increase sensitivity, some of these junctions such as silicon carbide were "biased" by connecting a battery and potentiometer across them to provide a small constant forward voltage across the junction.

The oxide layers that form on many ordinary metal surfaces have semiconducting properties, and detectors for crystal radios have been improvised from a variety of everyday objects such as rusty needles and corroded pennies. The *foxhole radio* was a crystal radio receiver improvised by soldiers during World War 2 without access to conventional sets. It used a razor blade and a safety pin or lead pencil to form a demodulating junction. Much patience was required to find an active detecting site on the blade. Stray rectifying junctions between metal parts of radio transmitter installations are still a source for interference, because they can radiate harmonics of the transmitter frequency.

## History



The modern circuit symbol for a diode originated as a schematic drawing of a cat's whisker detector.

Unlike modern radio stations that transmit audio (sound), the first radio transmitters during the first three decades of radio (1887-1917) transmitted information by telegraphy, turning the transmitter on and off with a switch called a telegraph key to spell out messages in Morse code, consisting of "dots" and "dashes". So early radio receiving apparatus merely had to detect the presence or absence of the radio signal, not convert it into audio. The device that did this was called a detector. The crystal detector was the most successful of many detector devices that were used in the early days of radio. It replaced electrolytic, magnetic, and particularly coherer detectors in radio receivers around 1906. Later, when AM radio transmission was developed to transmit sound, around World War 1, crystal detectors proved able to receive this as well.

The "unilateral conduction" of crystals, as it was then called, was discovered by Ferdinand Braun, a German physicist in 1874 at the University of Würzburg, before radio had been invented. Based on this work G.W. Pickard developed the cat's whisker diode using a silicon crystal, which was patented in 1906. However, Bengali scientist Jagadish Chandra Bose was the first to use a crystal to detect radio waves, in his experiments with microwaves in 1894, applying for a patent on a galena detector in 1901.

When these devices were in common use, more advanced proprietary versions of "permanent" detectors were developed, many of them by G. W. Pickard, who tested more than 30,000 combinations of crystal and wire contacts. One consisted of various combinations of pairs of different crystals such as Zincite touching Bornite or Chalcopyrite, in fairly heavily spring-loaded contact. Pickard named this the **Perikon detector**, from "**PER**fect p**I**c**K**ard c**ON**tact". Other detectors patented by Pickard included the common crystal iron pyrite. Pickard has the distinction of having brought silicon into use as a detector, patenting it in 1906. At nearly the same time, General Henry Harrison Chase Dunwoody patented the use of the silicon carbide (carborundum) detector, an artificial substance created accidentally during attempts by Edward Acheson to create diamonds.

Unamplified radio receivers, most of which were crystal radios, were the only way to receive radio signals during most of the wireless telegraphy era, which ended around 1920. Mineral detectors were largely superseded by vacuum tubes, invented in 1906, although the expense of tube receivers meant that full replacement took several decades. By the 1920s crystal radios were relegated to use by hobbyists and youth groups.

A crystal detector in commercial form from the 1960's

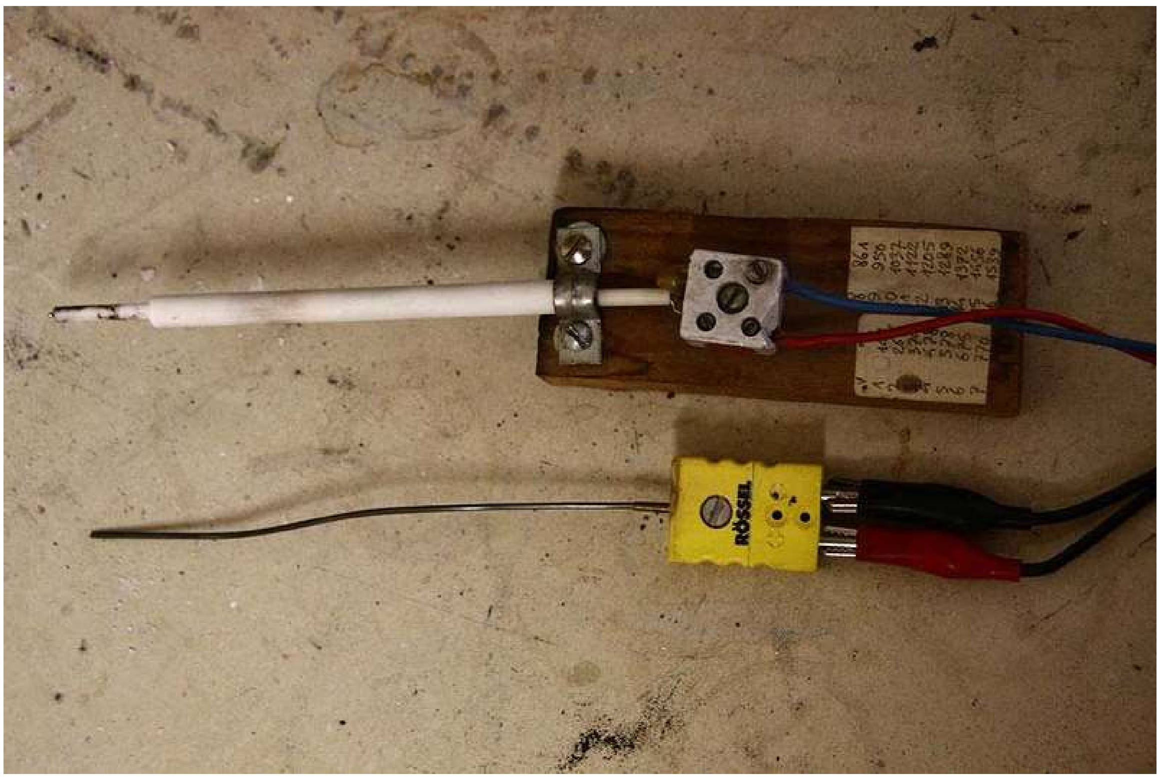
The point-contact semiconductor detector was subsequently resurrected around World War 2 because of the military requirement for microwave radar detectors. Vacuum tube detectors do not work at microwave frequencies. The small area of the point contact minimized minority carrier storage and capacitance, making these diodes fast enough to function at radar frequencies. Silicon and germanium point contact diodes were developed. Wartime research on PN junctions in crystals paved the way for the invention of the transistor in 1956. The first transistors also used cat's whisker contacts.

The germanium diodes that became widely available after the war proved to be as sensitive as galena and didn't require any adjustment, so they replaced cat's whisker detectors in crystal radios, largely putting an end to the manufacture of this antique radio component. Although cat's whisker detectors are obsolete, modern point-contact silicon detectors are still commercially produced. Thus the point contact method used to make these first semiconductor diodes 100 years ago is still being used today.

Copies of original cat's-whisker detectors are still manufactured and sold, for antique radio hobbyists.

# Chapter- 5

# Sensors



Thermocouple sensor for high temperature measurement

A *sensor*; is a device that measures a physical quantity and converts it into a signal which can be read by an observer or by an instrument. For example, a mercury-in-glass thermometer converts the measured temperature into expansion and contraction of a liquid which can be read on a calibrated glass tube. A thermocouple converts temperature to an output voltage which can be read by a voltmeter. For accuracy, most sensors are calibrated against known standards.

## Use

Sensors are used in everyday objects such as touch-sensitive elevator buttons (tactile sensor) and lamps which dim or brighten by touching the base. There are also innumerable applications for sensors of which most people are never aware. Applications include cars, machines, aerospace, medicine, manufacturing and robotics.

A sensor is a device which receives and responds to a signal. A sensor's sensitivity indicates how much the sensor's output changes when the measured quantity changes. For instance, if the mercury in a thermometer moves 1 cm when the temperature changes by 1 °C, the sensitivity is 1 cm/°C (it is basically the slope Dy/Dx assuming a linear characteristic). Sensors that measure very small changes must have very high sensitivities. Sensors also have an impact on what they measure; for instance, a room temperature thermometer inserted into a hot cup of liquid cools the liquid while the liquid heats the thermometer. Sensors need to be designed to have a small effect on what is measured, making the sensor smaller often improves this and may introduce other advantages. Technological progress allows more and more sensors to be manufactured on a microscopic scale as microsensors using MEMS technology. In most cases, a microsensor reaches a significantly higher speed and sensitivity compared with macroscopic approaches.

# Classification of measurement errors

A good sensor obeys the following rules:

- Is sensitive to the measured property
- Is insensitive to any other property likely to be encountered in its application
- Does not influence the measured property

Ideal sensors are designed to be linear or linear to some simple mathematical fuction of the measurement, typically logarithmic. The output signal of such a sensor is linearly proportional to the value or simple function of the measured property. The sensitivity is then defined as the ratio between output signal and measured property. For example, if a sensor measures temperature and has a voltage output, the sensitivity is a constant with the unit [V/K]; this sensor is linear because the ratio is constant at all points of measurement.

### Sensor deviations

If the sensor is not ideal, several types of deviations can be observed:

- The sensitivity may in practice differ from the value specified. This is called a sensitivity error, but the sensor is still linear.
- Since the range of the output signal is always limited, the output signal will eventually reach a minimum or maximum when the measured property exceeds the limits. The full scale range defines the maximum and minimum values of the measured property.
- If the output signal is not zero when the measured property is zero, the sensor has an offset or bias. This is defined as the output of the sensor at zero input.

- If the sensitivity is not constant over the range of the sensor, this is called nonlinearity. Usually this is defined by the amount the output differs from ideal behavior over the full range of the sensor, often noted as a percentage of the full range.
- If the deviation is caused by a rapid change of the measured property over time, there is a dynamic error. Often, this behaviour is described with a bode plot showing sensitivity error and phase shift as function of the frequency of a periodic input signal.
- If the output signal slowly changes independent of the measured property, this is defined as drift (telecommunication).
- Long term drift usually indicates a slow degradation of sensor properties over a long period of time.
- Noise is a random deviation of the signal that varies in time.
- Hysteresis is an error caused by when the measured property reverses direction, but there is some finite lag in time for the sensor to respond, creating a different offset error in one direction than in the other.
- If the sensor has a digital output, the output is essentially an approximation of the measured property. The approximation error is also called digitization error.
- If the signal is monitored digitally, limitation of the sampling frequency also can cause a dynamic error, or if the variable or added noise noise changes periodically at a frequency near a multiple of the sampling rate may induce aliasing errors.
- The sensor may to some extent be sensitive to properties other than the property being measured. For example, most sensors are influenced by the temperature of their environment.

All these deviations can be classified as systematic errors or random errors. Systematic errors can sometimes be compensated for by means of some kind of calibration strategy. Noise is a random error that can be reduced by signal processing, such as filtering, usually at the expense of the dynamic behaviour of the sensor.

### Resolution

The resolution of a sensor is the smallest change it can detect in the quantity that it is measuring. Often in a digital display, the least significant digit will fluctuate, indicating that changes of that magnitude are only just resolved. The resolution is related to the precision with which the measurement is made. For example, a scanning tunneling probe (a fine tip near a surface collects an electron tunnelling current) can resolve atoms and molecules.

## Types

This is a list of sensors sorted by sensor type.

## Acoustic, sound, vibration

- Geophone
- Hydrophone
- Lace Sensor a guitar pickup
- Microphone
- Seismometer
- Accelerometer

## Automotive, transportation

- Air-fuel ratio meter
- Crank sensor
- Curb feeler, used to warn driver of curbs
- Defect detector, used on railroads to detect axle and signal problems in passing trains
- MAP sensor, Manifold Absolute Pressure, used in regulating fuel metering.
- Parking sensors, used to alert the driver of unseen obstacles during parking manoeuvres
- Radar gun, used to detect the speed of other objects
- Speedometer, used measure the instantaneous speed of a land vehicle
- Speed sensor, used to detect the speed of an object
- Throttle position sensor, used to monitor the position of the throttle in an internal combustion engine
- Variable reluctance sensor, used to measure position and speed of moving metal components
- Water sensor or water-in-fuel sensor, used to indicate the presence of water in fuel
- Wheel speed sensor, used for reading the speed of a vehicle's wheel rotation

## Chemical

- Breathalyzer
- Carbon dioxide sensor
- Carbon monoxide detector
- Catalytic bead sensor
- Chemical field-effect transistor
- Electrochemical gas sensor
- Electronic nose
- Electrolyte–insulator–semiconductor sensor
- Hydrogen sensor
- Hydrogen sulfide sensor
- Infrared point sensor
- Ion-selective electrode
- Nondispersive infrared sensor
- Microwave chemistry sensor

- Nitrogen oxide sensor
- Olfactometer
- Optode
- Oxygen sensor
- Pellistor
- pH glass electrode
- Potentiometric sensor
- Redox electrode
- Smoke detector
- Zinc oxide nanorod sensor

## Electric current, electric potential, magnetic, radio

- Ammeter
- Current sensor
- Galvanometer
- Hall effect sensor
- Hall probe
- Leaf electroscope
- Magnetic anomaly detector
- Magnetometer
- Metal detector
- Multimeter
- Ohmmeter
- Radio direction finder
- Telescope
- Voltmeter
- Voltage detector
- Watt-hour meter

## Environment, weather, moisture, humidity

- Bedwetting alarm
- Dew warning
- Fish counter
- Gas detector
- Hook gauge evaporimeter
- Hygrometer
- Leaf sensor
- Pyranometer
- Pyrgeometer
- Psychrometer
- Rain gauge
- Rain sensor
- Seismometers

- Snow gauge
- Soil moisture sensor
- Stream gauge
- Tide gauge

## Flow, fluid velocity

- Air flow meter
- Anemometer
- Flow sensor
- Gas meter
- Mass flow sensor
- Water meter

## Ionising radiation, subatomic particles

- Bubble chamber
- Cloud chamber
- Geiger counter
- Neutron detection
- Particle detector
- Scintillation counter
- Scintillator
- Wire chamber

## Navigation instruments

- Air speed indicator
- Altimeter
- Attitude indicator
- Depth gauge
- Fluxgate compass
- Gyroscope
- Inertial reference unit
- Magnetic compass
- MHD sensor
- Ring laser gyroscope
- Turn coordinator
- Variometer
- Vibrating structure gyroscope
- Yaw rate sensor

# Position, angle, displacement, distance, speed, acceleration

- Accelerometer
- Capacitive displacement sensor
- Free fall sensor
- Gravimeter
- Inclinometer
- Laser rangefinder
- Linear encoder
- Linear variable differential transformer (LVDT)
- Liquid capacitive inclinometers
- Odometer
- Piezoelectric accelerometer
- Position sensor
- Rotary encoder
- Rotary variable differential transformer
- Selsyn
- Sudden Motion Sensor
- Tilt sensor
- Tachometer
- Ultrasonic thickness gauge

# Optical, light, imaging

- Charge-coupled device
- Colorimeter
- Contact image sensor
- Electro-optical sensor
- Flame detector
- Infra-red sensor
- LED as light sensor
- Nichols radiometer
- Fiber optic sensors
- Photodetector
- Photodiode
- Photomultiplier tubes
- Phototransistor
- Photoelectric sensor
- Photoionization detector
- Photomultiplier
- Photoresistor
- Photoswitch
- Phototube
- Proximity sensor

- Scintillometer
- Shack-Hartmann
- Wavefront sensor

## Pressure

- Barograph
- Barometer
- Boost gauge
- Bourdon gauge
- Hot filament ionization gauge
- Ionization gauge
- McLeod gauge
- Oscillating U-tube
- Permanent Downhole Gauge
- Pirani gauge
- Pressure sensor
- Pressure gauge
- Tactile sensor
- Time pressure gauge

## Force, density, level

- Bhangmeter
- Hydrometer
- Force gauge
- Level sensor
- Load cell
- Magnetic level gauge
- Nuclear density gauge
- Piezoelectric sensor
- Strain gauge
- Torque sensor
- Viscometer

## Thermal, heat, temperature

- Bolometer
- Calorimeter
- Exhaust gas temperature gauge
- Gardon gauge
- Heat flux sensor
- Infrared thermometer
- Microbolometer
- Microwave radiometer

- Net radiometer
- Resistance temperature detector
- Resistance thermometer
- Silicon bandgap temperature sensor
- Temperature gauge
- Thermistor
- Thermocouple
- Thermometer

## Proximity, presence

- Alarm sensor
- Motion detector
- Occupancy sensor
- Passive infrared sensor
- Reed switch
- Stud finder
- Triangulation sensor
- Touch switch
- Wired glove

## Sensor technology

- Active pixel sensor
- Machine vision
- Biochip
- Biosensor
- Capacitance probe
- Catadioptric sensor
- Carbon paste electrode
- Displacement receiver
- Electromechanical film
- Electro-optical sensor
- Fabry–Pérot interferometer
- Image sensor
- Inductive sensor
- Intelligent sensor
- Lab-on-a-chip
- Leaf sensor
- Micro-sensor arrays
- RADAR
- Sensor array
- Sensor grid
- Sensor node
- Soft sensor

- SONAR
- Staring array
- Transducer
- Ultrasonic sensor
- Video sensor
- Visual sensor network
- Wheatstone bridge
- Photoelasticity

# Sensors in Nature

All living organisms contain biological sensors with functions similar to those of the mechanical devices described. Most of these are specialized cells that are sensitive to:

- Light, motion, temperature, magnetic fields, gravity, humidity, vibration, pressure, electrical fields, sound, and other physical aspects of the external environment
- Physical aspects of the internal environment, such as stretch, motion of the organism, and position of appendages (proprioception)
- Environmental molecules, including toxins, nutrients, and pheromones
- Estimation of biomolecules interaction and some kinetics parameters
- Internal metabolic milieu, such as glucose level, oxygen level, or osmolality
- Internal signal molecules, such as hormones, neurotransmitters, and cytokines
- Differences between proteins of the organism itself and of the environment or alien creatures



# Biosensor

A **biosensor** is an analytical device for the detection of an analyte that combines a biological component with a physicochemical detector component.

It consists of 3 parts:

- the *sensitive biological element* (biological material (eg. tissue, microorganisms, organelles, cell receptors, enzymes, antibodies, nucleic acids, etc), a biologically derived material or biomimic) The sensitive elements can be created by biological engineering.
- the *transducer* or the *detector element* (works in a physicochemical way; optical, piezoelectric, electrochemical, etc.) that transforms the signal resulting from the interaction of the analyte with the biological element into another signal (i.e., transducers) that can be more easily measured and quantified;
- associated electronics or signal processors that are primarily responsible for the display of the results in a user-friendly way.. This sometimes accounts for the most expensive part of the sensor device, however it is possible to generate a user friendly display that includes transducer and sensitive element.

A common example of a commercial biosensor is the blood glucose biosensor, which uses the enzyme glucose oxidase to break blood glucose down. In doing so it first oxidizes glucose and uses two electrons to reduce the FAD (a component of the enzyme) to FADH2. This in turn is oxidized by the electrode (accepting two electrons from the electrode) in a number of steps. The resulting current is a measure of the concentration of glucose. In this case, the electrode is the transducer and the enzyme is the biologically active component.

Recently, arrays of many different detector molecules have been applied in so called electronic nose devices, where the pattern of response from the detectors is used to fingerprint a substance. Current commercial electronic noses, however, do not use biological elements.

A canary in a cage, as used by miners to warn of gas, could be considered a biosensor. Many of today's biosensor applications are similar, in that they use organisms which respond to toxic substances at a much lower concentrations than humans can detect to warn of the presence of the toxin. Such devices can be used in environmental monitoring, trace gas detection and in water treatment facilities.

# Principles of Detection

## Photometric

Many optical biosensors based on the phenomenon of surface plasmon resonance are evanescent wave techniques. This utilises a property of gold and other materials; specifically that a thin layer of gold on a high refractive index glass surface can absorb laser light, producing electron waves (surface plasmons) on the gold surface. This occurs only at a specific angle and wavelength of incident light and is highly dependent on the surface of the gold, such that binding of a target analyte to a receptor on the gold surface produces a measurable signal.

Surface plasmon resonance sensors operate using a sensor chip consisting of a plastic cassette supporting a glass plate, one side of which is coated with a microscopic layer of gold. This side contacts the optical detection apparatus of the instrument. The opposite side is then contacted with a microfluidic flow system. The contact with the flow system creates channels across which reagents can be passed in solution. This side of the glass sensor chip can be modified in a number of ways, to allow easy attachment of molecules of interest. Normally it is coated in carboxymethyl dextran or similar compound.

Light of a fixed wavelength is reflected off the gold side of the chip at the angle of total internal reflection, and detected inside the instrument. This induces the evanescent wave to penetrate through the glass plate and some distance into the liquid flowing over the surface.

The refractive index at the flow side of the chip surface has a direct influence on the behaviour of the light reflected off the gold side. Binding to the flow side of the chip has an effect on the refractive index and in this way biological interactions can be measured to a high degree of sensitivity with some sort of energy.

Other evanescent wave biosensors have been commercialised using waveguides where the propagation constant through the waveguide is changed by the absorption of molecules to the waveguide surface. One such example, Dual Polarisation Interferometry uses a buried waveguide as a reference against which the change in propagation constant is measured. Other configurations such as the Mach-Zehnder have reference arms lithographically defined on a substrate. Higher levels of integration can be achieved using resonator geometries where the resonant frequency of a ring resonator changes when molecules are absorbed.

Other optical biosensors are mainly based on changes in absorbance or fluorescence of an appropriate indicator compound and do not need a total internal reflection geometry. For example, a fully operational prototype device detecting casein in milk has been fabricated. The device is based on detecting changes in absorption of a gold layer. A widely used research tool, the micro-array, can also be considered a biosensor.

Biological biosensors often incorporate a genetically modified form of a native protein or enzyme. The protein is configured to detect a specific analyte and the ensuing signal is read by a detection instrument such as a fluorometer or luminometer. An example of a recently developed biosensor is one for detecting cytosolic concentration of the analyte cAMP (cyclic adenosine monophosphate), a second messenger involved in cellular signaling triggered by ligands interacting with receptors on the cell membrane. Similar systems have been created to study cellular responses to native ligands or xenobiotics (toxins or small molecule inhibitors). Such "assays" are commonly used in drug discovery development by pharmaceutical and biotechnology companies. Most cAMP assays in current use require lysis of the cells prior to measurement of cAMP. A live-cell biosensor for cAMP can be used in non-lysed cells with the additional advantage of multiple reads to study the kinetics of receptor response.

## Electrochemical

Electrochemical biosensors are normally based on enzymatic catalysis of a reaction that produces or consumes electrons (such enzymes are rightly called redox enzymes). The sensor substrate usually contains three electrodes; a reference electrode, a working electrode and a sink electrode. An auxiliary electrode (also known as a counter electrode) may also be present as an ion source. The target analyte is involved in the reaction that takes place on the active electrode surface, and the ions produced create a potential which is subtracted from that of the reference electrode to give a signal. We can either measure the current (rate of flow of electrons is now proportional to the analyte concentration) at a fixed potential or the potential can be measured at zero current (this gives a logarithmic response). Note that potential of the working or active electrode is space charge sensitive and this is often used. Further, the label-free and direct electrical detection of small peptides and proteins is possible by their intrinsic charges using biofunctionalized ion-sensitive field-effect transistors.

Another example, the potentiometric biosensor, works contrary to the current understanding of its ability. Such biosensors are screenprinted, conducting polymer coated, open circuit potential biosensors based on conjugated polymers immunoassays. They have only two electrodes and are extremely sensitive and robust. They enable the detection of analytes at levels previously only achievable by HPLC and LC/MS and without rigorous sample preparation. The signal is produced by electrochemical and physical changes in the conducting polymer layer due to changes occurring at the surface of the sensor. Such changes can be attributed to ionic strength, pH, hydration and redox reactions, the latter due to the enzyme label turning over a substrate().

### Others

Piezoelectric sensors utilise crystals which undergo an elastic deformation when an electrical potential is applied to them. An alternating potential (A.C.) produces a standing wave in the crystal at a characteristic frequency. This frequency is highly dependent on the elastic properties of the crystal, such that if a crystal is coated with a biological recognition element the binding of a (large) target analyte to a receptor will produce a change in the resonance frequency, which gives a binding signal. In a mode that uses surface acoustic waves (SAW), the sensitivity is greatly increased. This is a specialised application of the Quartz crystal microbalance as a biosensor.

Thermometric and magnetic based biosensors are rare.

## Applications

There are many potential applications of biosensors of various types. The main requirements for a biosensor approach to be valuable in terms of research and commercial applications are the identification of a target molecule, availability of a suitable biological recognition element, and the potential for disposable portable detection systems to be preferred to sensitive laboratory-based techniques in some situations. Some examples are given below:

- Glucose monitoring in diabetes patients ←**historical market driver**
- Other medical health related targets
- Environmental applications e.g. the detection of pesticides and river water contaminants
- Remote sensing of airborne bacteria e.g. in counter-bioterrorist activities
- Detection of pathogens
- Determining levels of toxic substances before and after bioremediation
- Detection and determining of organophosphate
- Routine analytical measurement of folic acid, biotin, vitamin B12 and pantothenic acid as an alternative to microbiological assay
- Determination of drug residues in food, such as antibiotics and growth promoters, particularly meat and honey.
- Drug discovery and evaluation of biological activity of new compounds.
- Protein engineering in biosensors

- Detection of toxic metabolites such as mycotoxins

### Glucose monitoring

Commercially available gluocose monitors rely on amperometric sensing of glucose by means of glucose oxidase, which oxidises glucose producing hydrogen peroxide which is detected by the electrode. To overcome the limitation of ameperometric sensors, a flurry of research is present into novel sensing methods, such as fluorescent glucose biosensors.

## Biosensors in food analysis

There are several applications of biosensors in food analysis. In food industry optic coated with antibodies are commonly used to detect pathogens and food toxins. The light system in these biosensors has been fluorescence, since this type of optical measurement can greatly amplify the signal.

A range of immuno- and ligand-binding assays for the detection and measurement of small molecules such as water-soluble vitamins and chemical contaminants (drug residues) such as sulfonamides and Beta-agonists have been developed for use on SPR based sensor systems, often adapted from existing ELISA or other immunological assay. These are in widespread use across the food industry.

## Surface Attachment of the biological elements

An important part in a biosensor is to attach the biological elements (small molecules/protein/cells) to the surface of the sensor (be it metal, polymer or glass). The simplest way is to functionalize the surface in order to coat it with the biological elements. This can be done by polylysine, aminosilane, epoxysilane or nitrocellulose in the case of silicon chips/silica glass. Subsequently the bound biological agent may be for example fixed by Layer by layer depositation of alternatively charged polymer coatings Alternatively three dimensional lattices (hydrogel/xerogel) can be used to chemically or physically entrap these (where by chemically entraped it is meant that the biological element is kept in place by a strong bond, while physically they are kept in place being unable to pass through the pores of the gel matrix). The most commonly used hydrogel is sol-gel, a glassy silica generated by polymerization of silicate monomers (added as tetra alkyl orthosilicates, such as TMOS or TEOS) in the presence of the biological elements (along with other stabilizing polymers, such as PEG) in the case of physical entrapment. Another group of hydrogels, which set under conditions suitable for cells or protein, are acrylate hydrogel, which polymerize upon radical initiation. One type of radical initiator is a peroxide radical, typically generated by combining a persulfate with TEMED (Polyacrylamide gel are also commonly commonly used for protein electrophoresis), alternatively light can be used in combination with a photoinitiator, such as DMPA (2,2-dimethoxy-2-phenylacetophenone). Smart materials that mimic the biological components of a sensor can also be classified as biosensors using only the active or catalytic site or analogous configurations of a biomolecule.

## Chapter- 6

# Microphone


A Neumann U87 condenser microphone with shock mount

A **microphone** (colloquially called a **mic** or **mike**; both pronounced /ˈmaɪk/) is an acoustic-to-electric transducer or sensor that converts sound into an electrical signal. In 1876, Emile Berliner invented the first microphone used as a telephone voice transmitter. Microphones are used in many applications such as telephones, tape recorders, karaoke systems, hearing aids, motion picture production, live and recorded audio engineering, FRS radios, megaphones, in radio and television broadcasting and in computers for recording voice, speech recognition, VoIP, and for non-acoustic purposes such as ultrasonic checking or knock sensors.

Most microphones today use electromagnetic induction (dynamic microphone), capacitance change (condenser microphone), piezoelectric generation, or light modulation to produce an electrical voltage signal from mechanical vibration.

# Varieties

The sensitive transducer element of a microphone is called its *element* or *capsule.* A complete microphone also includes a housing, some means of bringing the signal from the element to other equipment, and often an electronic circuit to adapt the output of the capsule to the equipment being driven. Microphones are referred to by their transducer principle, such as condenser, dynamic, etc., and by their directional characteristics. Sometimes other characteristics such as diaphragm size, intended use or orientation of the principal sound input to the principal axis (end- or side-address) of the microphone are used to describe the microphone.

## Condenser microphone

Inside the Oktava 319 condenser microphone

The **condenser microphone**, invented at Bell Labs in 1916 by E. C. Wente is also called a **capacitor microphone** or **electrostatic microphone**. Here, the diaphragm acts as one plate of a capacitor, and the vibrations produce changes in the distance between the plates. There are two types, depending on the method of extracting the audio signal from the transducer: DC-biased and radio frequency (RF) or high frequency (HF) condenser microphones. With a DC-biased microphone, the plates are biased with a fixed charge (*Q*). The voltage maintained across the capacitor plates changes with the vibrations in the air, according to the capacitance equation (C = Q / V), where Q = charge in coulombs, C = capacitance in farads and V = potential difference in volts. The capacitance of the plates is inversely proportional to the distance between them for a parallel-plate capacitor. The assembly of fixed and movable plates is called an "element" or "capsule."

A nearly constant charge is maintained on the capacitor. As the capacitance changes, the charge across the capacitor does change very slightly, but at audible frequencies it is sensibly constant. The capacitance of the capsule (around 5 to 100 pF) and the value of the bias resistor (100 megohms to tens of gigohms) form a filter that is high-pass for the audio signal, and low-pass for the bias voltage. Note that the time constant of an RC circuit equals the product of the resistance and capacitance.

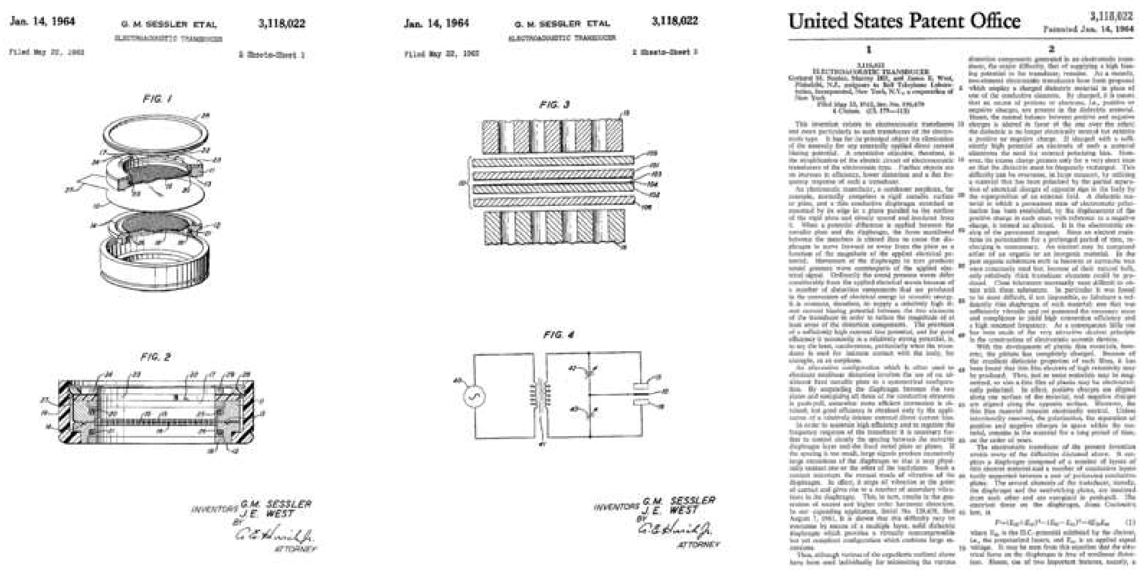
Within the time-frame of the capacitance change (as much as 50 ms at 20 Hz audio signal), the charge is practically constant and the voltage across the capacitor changes instantaneously to reflect the change in capacitance. The voltage across the capacitor varies above and below the bias voltage. The voltage difference between the bias and the capacitor is seen across the series resistor. The voltage across the resistor is amplified for performance or recording.


AKG C451B small-diaphragm condenser microphone

RF condenser microphones use a comparatively low RF voltage, generated by a low-noise oscillator. The oscillator may either be amplitude modulated by the capacitance changes produced by the sound waves moving the capsule diaphragm, or the capsule may be part of a resonant circuit that modulates the frequency of the oscillator signal. Demodulation yields a low-noise audio frequency signal with a very low source impedance. The absence of a high bias voltage permits the use of a diaphragm with looser tension, which may be used to achieve wider frequency response due to higher compliance. The RF biasing process results in a lower electrical impedance capsule, a useful by-product of which is that RF condenser microphones can be operated in damp weather conditions that could create problems in DC-biased microphones with contaminated insulating surfaces. The Sennheiser "MKH" series of microphones use the RF biasing technique.

Condenser microphones span the range from telephone transmitters through inexpensive karaoke microphones to high-fidelity recording microphones. They generally produce a high-quality audio signal and are now the popular choice in laboratory and studio recording applications. The inherent suitability of this technology is due to the very small mass that must be moved by the incident sound wave, unlike other microphone types that require the sound wave to do more work. They require a power source, provided either via microphone outputs as phantom power or from a small battery. Power is necessary for establishing the capacitor plate voltage, and is also needed to power the microphone electronics (impedance conversion in the case of electret and DC-polarized microphones, demodulation or detection in the case of RF/HF microphones). Condenser microphones are also available with two diaphragms that can be electrically connected to provide a range of polar patterns (see below), such as cardioid, omnidirectional, and figure-eight. It is also possible to vary the pattern continuously with some microphones, for example the Røde NT2000 or CAD M179.

## Electret condenser microphone




First patent on foil electret microphone

An electret microphone is a relatively new type of capacitor microphone invented at Bell laboratories in 1962 by Gerhard Sessler and Jim West. The externally applied charge described above under condenser microphones is replaced by a permanent charge in an electret material. An electret is a ferroelectric material that has been permanently electrically charged or *polarized*. The name comes from *electr*ostatic and magn*et*; a static charge is embedded in an electret by alignment of the static charges in the material, much the way a magnet is made by aligning the magnetic domains in a piece of iron.

Due to their good performance and ease of manufacture, hence low cost, the vast majority of microphones made today are electret microphones; a semiconductor manufacturer estimates annual production at over one billion units. Nearly all cell-phone, computer, PDA and headset microphones are electret types. They are used in many applications, from high-quality recording and lavalier use to built-in microphones in small sound recording devices and telephones. Though electret microphones were once considered low quality, the best ones can now rival traditional condenser microphones in every respect and can even offer the long-term stability and ultra-flat response needed for a measurement microphone. Unlike other capacitor microphones, they require no polarizing voltage, but often contain an integrated preamplifier that does require power (often incorrectly called polarizing power or bias). This preamplifier is frequently phantom powered in sound reinforcement and studio applications. Microphones designed for personal computer (PC) use, sometimes called multimedia microphones, use a stereo 3.5 mm plug (though a mono source) with the ring receiving power via a resistor from (normally) a 5 V supply in the computer; unfortunately, a number of incompatible dynamic microphones are fitted with 3.5 mm plugs too. While few electret microphones rival the best DC-polarized units in terms of noise level, this is not due to any inherent limitation of the electret. Rather, mass production techniques needed to produce microphones cheaply don't lend themselves to the precision needed to produce the highest quality microphones, due to the tight tolerances required in internal dimensions. These tolerances are the same for all condenser microphones, whether the DC, RF or electret technology is used.

### Dynamic microphone

Patti Smith singing into a Shure SM58 (dynamic cardioid type) microphone

**Dynamic microphones** work via electromagnetic induction. They are robust, relatively inexpensive and resistant to moisture. This, coupled with their potentially high gain before feedback makes them ideal for on-stage use.

Moving-coil microphones use the same dynamic principle as in a loudspeaker, only reversed. A small movable induction coil, positioned in the magnetic field of a permanent magnet, is attached to the diaphragm. When sound enters through the windscreen of the microphone, the sound wave moves the diaphragm. When the diaphragm vibrates, the coil moves in the magnetic field, producing a varying current in the coil through electromagnetic induction. A single dynamic membrane does not respond linearly to all audio frequencies. Some microphones for this reason utilize multiple membranes for the different parts of the audio spectrum and then combine the resulting signals. Combining the multiple signals correctly is difficult and designs that do this are rare and tend to be expensive. There are on the other hand several designs that are more specifically aimed towards isolated parts of the audio spectrum. The AKG D 112, for example, is designed for bass response rather than treble. In audio engineering several kinds of microphones are often used at the same time to get the best result.

## Ribbon Microphone


Edmund Lowe using a ribbon microphone

Ribbon microphones use a thin, usually corrugated metal ribbon suspended in a magnetic field. The ribbon is electrically connected to the microphone's output, and its vibration within the magnetic field generates the electrical signal. Ribbon microphones are similar to moving coil microphones in the sense that both produce sound by means of magnetic induction. Basic ribbon microphones detect sound in a bi-directional (also called figure-eight) pattern because the ribbon, which is open to sound both front and back, responds to the pressure gradient rather than the sound pressure. Though the symmetrical front and rear pickup can be a nuisance in normal stereo recording, the high side rejection can be used to advantage by positioning a ribbon microphone horizontally, for example above cymbals, so that the rear lobe picks up only sound from the cymbals. Crossed figure 8, or Blumlein pair, stereo recording is gaining in popularity, and the figure 8 response of a ribbon microphone is ideal for that application.

Other directional patterns are produced by enclosing one side of the ribbon in an acoustic trap or baffle, allowing sound to reach only one side. The classic RCA Type 77-DX microphone has several externally adjustable positions of the internal baffle, allowing the selection of several response patterns ranging from "Figure-8" to "Unidirectional". Such older ribbon microphones, some of which still provide high quality sound reproduction, were once valued for this reason, but a good low-frequency response could only be obtained when the ribbon was suspended very loosely, which made them relatively fragile. Modern ribbon materials, including new nanomaterials have now been introduced that eliminate those concerns, and even improve the effective dynamic range of ribbon microphones at low frequencies. Protective wind screens can reduce the danger of damaging a vintage ribbon, and also reduce plosive artifacts in the recording. Properly designed wind screens produce negligible treble attenuation. In common with other classes of dynamic microphone, ribbon microphones don't require phantom power; in fact, this voltage can damage some older ribbon microphones. Some new modern ribbon microphone designs incorporate a preamplifier and, therefore, do require phantom power, and circuits of modern passive ribbon microphones, *i.e.*, those without the aforementioned preamplifier, are specifically designed to resist damage to the ribbon and transformer by phantom power. Also there are new ribbon materials available that are immune to wind blasts and phantom power.

## Carbon microphone

A carbon microphone, also known as a carbon button microphone (or sometimes just a button microphone), use a capsule or button containing carbon granules pressed between two metal plates like the Berliner and Edison microphones. A voltage is applied across the metal plates, causing a small current to flow through the carbon. One of the plates, the diaphragm, vibrates in sympathy with incident sound waves, applying a varying pressure to the carbon. The changing pressure deforms the granules, causing the contact area between each pair of adjacent granules to change, and this causes the electrical resistance of the mass of granules to change. The changes in resistance cause a corresponding change in the current flowing through the microphone, producing the electrical signal. Carbon microphones were once commonly used in telephones; they have extremely low-quality sound reproduction and a very limited frequency response range, but are very robust devices. The Boudet microphone, which used relatively large carbon balls, was similar to the granule carbon button microphones.

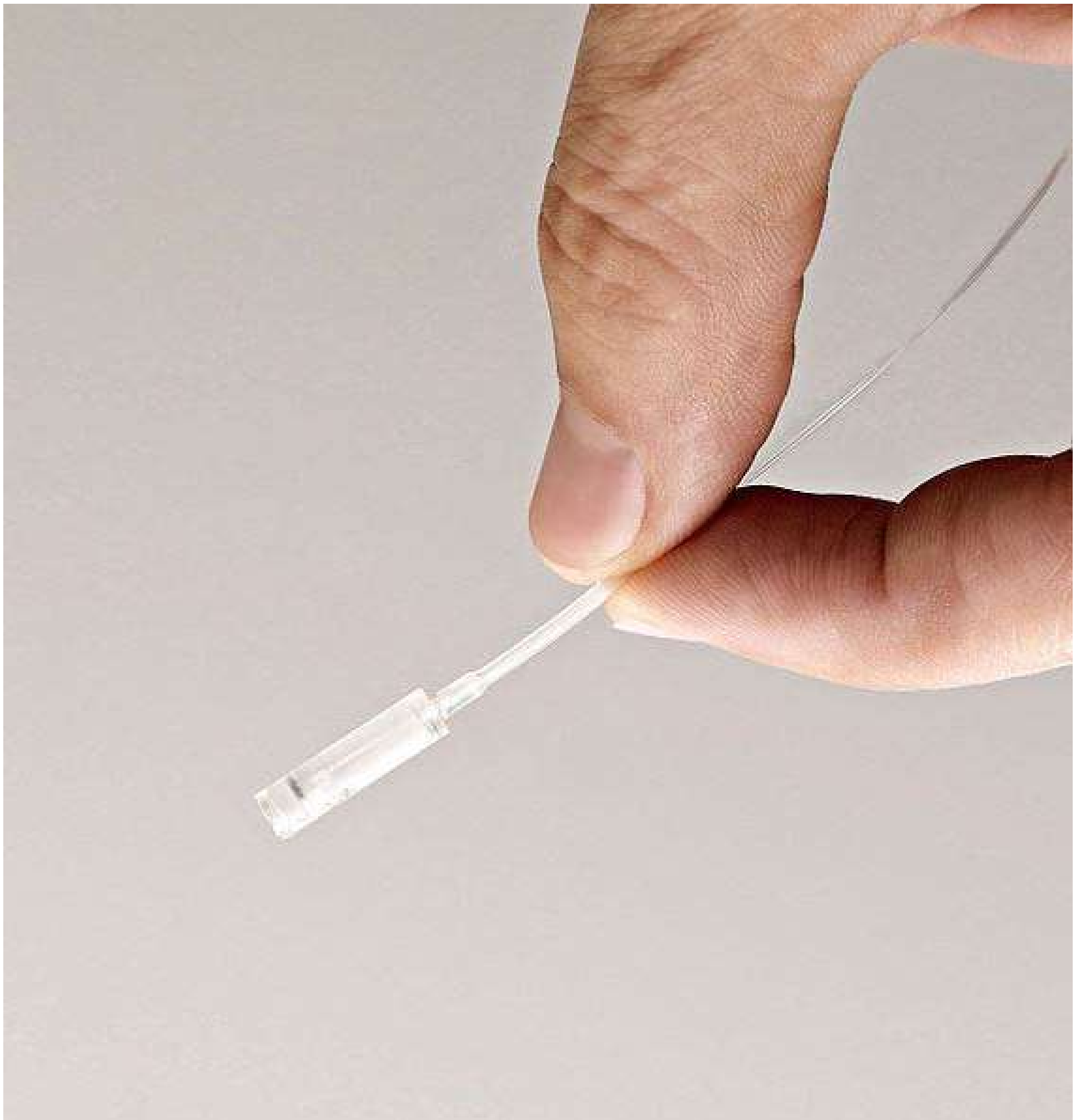
Unlike other microphone types, the carbon microphone can also be used as a type of amplifier, using a small amount of sound energy to control a larger amount of electrical energy. Carbon microphones found use as early telephone repeaters, making long distance phone calls possible in the era before vacuum tubes. These repeaters worked by mechanically coupling a magnetic telephone receiver to a carbon microphone: the faint signal from the receiver was transferred to the microphone, with a resulting stronger electrical signal to send down the line. One illustration of this amplifier effect was the oscillation caused by feedback, resulting in an audible squeal from the old "candlestick" telephone if its earphone was placed near the carbon microphone.

## Piezoelectric microphone

A **crystal microphone** or **piezo microphone** uses the phenomenon of piezoelectricity — the ability of some materials to produce a voltage when subjected to pressure — to convert vibrations into an electrical signal. An example of this is potassium sodium tartrate, which is a piezoelectric crystal that works as a transducer, both as a microphone and as a slimline loudspeaker component. Crystal microphones were once commonly supplied with vacuum tube (valve) equipment, such as domestic tape recorders. Their high output impedance matched the high input impedance (typically about 10 megohms) of the vacuum tube input stage well. They were difficult to match to early transistor equipment, and were quickly supplanted by dynamic microphones for a time, and later small electret condenser devices. The high impedance of the crystal microphone made it very susceptible to handling noise, both from the microphone itself and from the connecting cable.

Piezoelectric transducers are often used as contact microphones to amplify sound from acoustic musical instruments, to sense drum hits, for triggering electronic samples, and to record sound in challenging environments, such as underwater under high pressure. Saddle-mounted pickups on acoustic guitars are generally piezoelectric devices that contact the strings passing over the saddle. This type of microphone is different from magnetic coil pickups commonly visible on typical electric guitars, which use magnetic induction, rather than mechanical coupling, to pick up vibration.

## Fiber optic microphone



The Optoacoustics 1140 fiber optic microphone

A fiber optic microphone converts acoustic waves into electrical signals by sensing changes in light intensity, instead of sensing changes in capacitance or magnetic fields as with conventional microphones.

During operation, light from a laser source travels through an optical fiber to illuminate the surface of a tiny, sound-sensitive reflective diaphragm. Sound causes the diaphragm to vibrate, thereby minutely changing the intensity of the light it reflects. The modulated light is then transmitted over a second optical fiber to a photo detector, which transforms the intensity-modulated light into analog or digital audio for transmission or recording.

Fiber optic microphones possess high dynamic and frequency range, similar to the best high fidelity conventional microphones.

Fiber optic microphones do not react to or influence any electrical, magnetic, electrostatic or radioactive fields (this is called EMI/RFI immunity). The fiber optic microphone design is therefore ideal for use in areas where conventional microphones are ineffective or dangerous, such as inside industrial turbines or in magnetic resonance imaging (MRI) equipment environments.

Fiber optic microphones are robust, resistant to environmental changes in heat and moisture, and can be produced for any directionality or impedance matching. The distance between the microphone's light source and its photo detector may be up to several kilometers without need for any preamplifier and/or other electrical device, making fiber optic microphones suitable for industrial and surveillance acoustic monitoring.

Fiber optic microphones are used in very specific application areas such as for infrasound monitoring and noise-canceling. They have proven especially useful in medical applications, such as allowing radiologists, staff and patients within the powerful and noisy magnetic field to converse normally, inside the MRI suites as well as in remote control rooms.) Other uses include industrial equipment monitoring and sensing, audio calibration and measurement, high-fidelity recording and law enforcement.

## Laser microphone

Laser microphones are often portrayed in movies as spy gadgets. A laser beam is aimed at the surface of a window or other plane surface that is affected by sound. The slight vibrations of this surface displace the returned beam, causing it to trace the sound wave. The vibrating laser spot is then converted back to sound. In a more robust and expensive implementation, the returned light is split and fed to an interferometer, which detects movement of the surface. The former implementation is a tabletop experiment; the latter requires an extremely stable laser and precise optics.

A new type of laser microphone is a device that uses a laser beam and smoke or vapor to detect sound vibrations in free air. On 25 August 2009, U.S. patent 7,580,533 issued for a Particulate Flow Detection Microphone based on a laser-photocell pair with a moving stream of smoke or vapor in the laser beam's path. Sound pressure waves cause disturbances in the smoke that in turn cause variations in the amount of laser light reaching the photo detector. A prototype of the device was demonstrated at the 127th Audio Engineering Society convention in New York City from 9 through 12 October 2009.

## Liquid microphone

Early microphones did not produce intelligible speech, until Alexander Graham Bell made improvements including a variable resistance microphone/transmitter. Bell's liquid transmitter consisted of a metal cup filled with water with a small amount of sulfuric acid added. A sound wave caused the diaphragm to move, forcing a needle to move up and down in the water. The electrical resistance between the wire and the cup was then inversely proportional to the size of the water meniscus around the submerged needle. Elisha Gray filed a caveat for a version using a brass rod instead of the needle. Other minor variations and improvements were made to the liquid microphone by Majoranna, Chambers, Vanni, Sykes, and Elisha Gray, and one version was patented by Reginald Fessenden in 1903. These were the first working microphones, but they were not practical for commercial application. The famous first phone conversation between Bell and Watson took place using a liquid microphone.

## MEMS microphone

The MEMS (MicroElectrical-Mechanical System) microphone is also called a microphone chip or silicon microphone. The pressure-sensitive diaphragm is etched directly into a silicon chip by MEMS techniques, and is usually accompanied with integrated preamplifier. Most MEMS microphones are variants of the condenser microphone design. Often MEMS microphones have built in analog-to-digital converter (ADC) circuits on the same CMOS chip making the chip a digital microphone and so more readily integrated with modern digital products. Major manufacturers producing MEMS silicon microphones are Wolfson Microelectronics (WM7xxx), Analog Devices, Akustica (AKU200x), Infineon (SMM310 product), Knowles Electronics, Memstech (MSMx), NXP Semiconductors, Sonion MEMS, AAC Acoustic Technologies, and Omron.

## Speakers as microphones

A loudspeaker, a transducer that turns an electrical signal into sound waves, is the functional opposite of a microphone. Since a conventional speaker is constructed much like a dynamic microphone (with a diaphragm, coil and magnet), speakers can actually work "in reverse" as microphones. The result, though, is a microphone with poor quality, limited frequency response (particularly at the high end), and poor sensitivity. In practical use, speakers are sometimes used as microphones in applications where high quality and sensitivity are not needed such as intercoms, walkie-talkies or Video game voice chat peripherals, or when conventional microphones are in short supply.

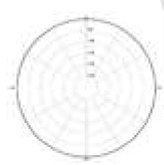
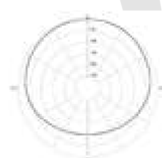
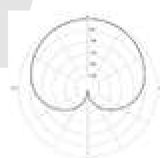
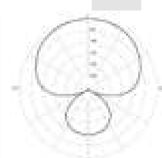
However, there is at least one other practical application of this principle: Using a medium-size woofer placed closely in front of a "kick" (bass drum) in a drum set to act as a microphone. The use of relatively large speakers to transduce low frequency sound sources, especially in music production, is becoming fairly common. A product example of this type of device is the Yamaha Subkick, a 6.5-inch (170 mm) woofer shock-mounted it into a 10" drum shell used in front of kick drums. Since a relatively massive membrane is unable to transduce high frequencies, placing a speaker in front of a kick drum is often ideal for reducing cymbal and snare bleed into the kick drum sound. Less commonly, microphones themselves can be used as speakers, almost always as tweeters. Microphones, however, are not designed to handle the power that speaker components are routinely required to cope with. One instance of such an application was the STC microphone-derived 4001 super-tweeter, which was successfully used in a number of high quality loudspeaker systems from the late 1960s to the mid-70s.

## Capsule design and directivity

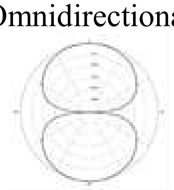
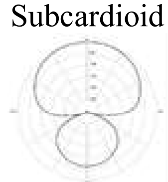
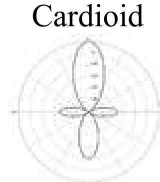
The inner elements of a microphone are the primary source of differences in directivity. A pressure microphone uses a diaphragm between a fixed internal volume of air and the environment, and responds uniformly to pressure from all directions, so it is said to be omnidirectional. A pressure-gradient microphone uses a diaphragm that is at least partially open on both sides. The pressure difference between the two sides produces its directional characteristics. Other elements such as the external shape of the microphone and external devices such as interference tubes can also alter a microphone's directional response. A pure pressure-gradient microphone is equally sensitive to sounds arriving from front or back, but insensitive to sounds arriving from the side because sound arriving at the front and back at the same time creates no gradient between the two. The characteristic directional pattern of a pure pressure-gradient microphone is like a figure-8. Other polar patterns are derived by creating a capsule that combines these two effects in different ways. The cardioid, for instance, features a partially closed backside, so its response is a combination of pressure and pressure-gradient characteristics.

## Microphone polar patterns

(Microphone facing top of page in diagram, parallel to page):

Omnidirectional Subcardioid Cardioid Supercardioid

Bi-directional or Figure of 8Hypercardioid Shotgun

A microphone's directionality or polar pattern indicates how sensitive it is to sounds arriving at different angles about its central axis. The polar patterns illustrated above represent the locus of points that produce the same signal level output in the microphone if a given sound pressure level is generated from that point. How the physical body of the microphone is oriented relative to the diagrams depends on the microphone design. For large-membrane microphones such as in the Oktava (pictured above), the upward direction in the polar diagram is usually perpendicular to the microphone body, commonly known as "side fire" or "side address". For small diaphragm microphones such as the Shure (also pictured above), it usually extends from the axis of the microphone commonly known as "end fire" or "top/end address".

Some microphone designs combine several principles in creating the desired polar pattern. This ranges from shielding (meaning diffraction/dissipation/absorption) by the housing itself to electronically combining dual membranes.

## Omnidirectional

An omnidirectional (or nondirectional) microphone's response is generally considered to be a perfect sphere in three dimensions. In the real world, this is not the case. As with directional microphones, the polar pattern for an "omnidirectional" microphone is a function of frequency. The body of the microphone is not infinitely small and, as a consequence, it tends to get in its own way with respect to sounds arriving from the rear, causing a slight flattening of the polar response. This flattening increases as the diameter of the microphone (assuming it's cylindrical) reaches the wavelength of the frequency in question. Therefore, the smallest diameter microphone gives the best omnidirectional characteristics at high frequencies.

The wavelength of sound at 10 kHz is little over an inch (3.4 cm) so the smallest measuring microphones are often 1/4" (6 mm) in diameter, which practically eliminates directionality even up to the highest frequencies. Omnidirectional microphones, unlike cardioids, do not employ resonant cavities as delays, and so can be considered the "purest" microphones in terms of low coloration; they add very little to the original sound. Being pressure-sensitive they can also have a very flat low-frequency response down to 20 Hz or below. Pressure-sensitive microphones also respond much less to wind noise and plosives than directional (velocity sensitive) microphones.

An example of a nondirectional microphone is the round black *eight ball*.

## Unidirectional

A unidirectional microphone is sensitive to sounds from only one direction. The diagram above illustrates a number of these patterns. The microphone faces upwards in each diagram. The sound intensity for a particular frequency is plotted for angles radially from 0 to 360°. (Professional diagrams show these scales and include multiple plots at different frequencies. The diagrams given here provide only an overview of typical pattern shapes, and their names.)

## Cardioids


US664A University Sound Dynamic Supercardioid Microphone

The most common unidirectional microphone is a cardioid microphone, so named because the sensitivity pattern is heart-shaped. A hyper-cardioid microphone is similar but with a tighter area of front sensitivity and a smaller lobe of rear sensitivity. A super-cardioid microphone is similar to a hyper-cardioid, except there is more front pickup and less rear pickup. These three patterns are commonly used as vocal or speech microphones, since they are good at rejecting sounds from other directions.

A cardioid microphone is effectively a superposition of an omnidirectional and a figure-8 microphone; for sound waves coming from the back, the negative signal from the figure-8 cancels the positive signal from the omnidirectional element, whereas for sound waves coming from the front, the two add to each other. A hypercardioid microphone is similar, but with a slightly larger figure-8 contribution. Since pressure gradient transducer microphones are directional, putting them very close to the sound source (at distances of a few centimeters) results in a bass boost. This is known as the proximity effect.

## Bi-directional

"Figure 8" or bi-directional microphones receive sound from both the front and back of the element. Most ribbon microphones are of this pattern.

## Shotgun

An Audio-Technica shotgun microphone

**Shotgun microphones** are the most highly directional. They have small lobes of sensitivity to the left, right, and rear but are significantly less sensitive to the side and rear than other directional microphones. This results from placing the element at the end of a tube with slots cut along the side; wave cancellation eliminates much of the off-axis sound. Due to the narrowness of their sensitivity area, shotgun microphones are commonly used on television and film sets, in stadiums, and for field recording of wildlife.

## Boundary or "PZM"

Several approaches have been developed for effectively using a microphone in less-than-ideal acoustic spaces, which often suffer from excessive reflections from one or more of the surfaces (boundaries) that make up the space. If the microphone is placed in, or very close to, one of these boundaries, the reflections from that surface are not sensed by the microphone. Initially this was done by placing an ordinary microphone adjacent to the surface, sometimes in a block of acoustically transparent foam. Sound engineers Ed Long and Ron Wickersham developed the concept of placing the diaphgram parallel to and facing the boundary. While the patent has expired, "Pressure Zone Microphone" and "PZM" are still active trademarks of Crown International, and the generic term "boundary microphone" is preferred. While a boundary microphone was initially implemented using an omnidirectional element, it is also possible to mount a directional microphone close enough to the surface to gain some of the benefits of this technique while retaining the directional properties of the element. Crown's trademark on this approach is "Phase Coherent Cardioid" or "PCC," but there are other makers who employ this technique as well.

## Application-specific designs

A lavalier microphone is made for hands-free operation. These small microphones are worn on the body. Originally, they were held in place with a lanyard worn around the neck, but more often they are fastened to clothing with a clip, pin, tape or magnet. The lavalier cord may be hidden by clothes and either run to an RF transmitter in a pocket or clipped to a belt (for mobile use), or run directly to the mixer (for stationary applications).

A wireless microphone transmits the audio as a radio or optical signal rather than via a cable. It usually sends its signal using a small FM radio transmitter to a nearby receiver connected to the sound system, but it can also use infrared waves if the transmitter and receiver are within sight of each other.

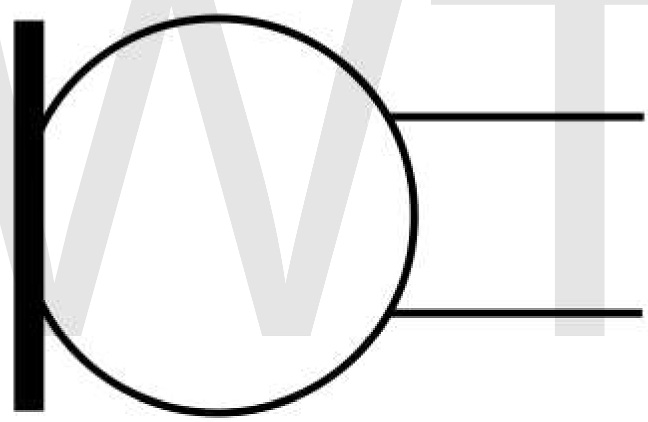
A contact microphone picks up vibrations directly from a solid surface or object, as opposed to sound vibrations carried through air. One use for this is to detect sounds of a very low level, such as those from small objects or insects. The microphone commonly consists of a magnetic (moving coil) transducer, contact plate and contact pin. The contact plate is placed directly on the vibrating part of a musical instrument or other surface, and the contact pin transfers vibrations to the coil. Contact microphones have been used to pick up the sound of a snail's heartbeat and the footsteps of ants. A portable version of this microphone has recently been developed. A throat microphone is a variant of the contact microphone that picks up speech directly from a person's throat, which it is strapped to. This lets the device be used in areas with ambient sounds that would otherwise make the speaker inaudible.

A parabolic microphone uses a parabolic reflector to collect and focus sound waves onto a microphone receiver, in much the same way that a parabolic antenna (e.g. satellite dish) does with radio waves. Typical uses of this microphone, which has unusually focused front sensitivity and can pick up sounds from many meters away, include nature recording, outdoor sporting events, eavesdropping, law enforcement, and even espionage. Parabolic microphones are not typically used for standard recording applications, because they tend to have poor low-frequency response as a side effect of their design.

A stereo microphone integrates two microphones in one unit to produce a stereophonic signal. A stereo microphone is often used for broadcast applications or field recording where it would be impractical to configure two separate condenser microphones in a classic X-Y configuration for stereophonic recording. Some such microphones have an adjustable angle of coverage between the two channels.

A noise-canceling microphone is a highly directional design intended for noisy environments. One such use is in aircraft cockpits where they are normally installed as boom microphones on headsets. Another use is on loud concert stages for vocalists. Many noise-canceling microphones combine signals received from two diaphragms that are in opposite electrical polarity or are processed electronically. In dual diaphragm designs, the main diaphragm is mounted closest to the intended source and the second is positioned farther away from the source so that it can pick up environmental sounds to be subtracted from the main diaphragm's signal. After the two signals have been combined, sounds other than the intended source are greatly reduced, substantially increasing intelligibility. Other noise-canceling designs use one diaphragm that is affected by ports open to the sides and rear of the microphone, with the sum being a 16 dB rejection of sounds that are farther away. One noise-canceling headset design using a single diaphragm has been used prominently by vocal artists such as Garth Brooks and Janet Jackson. A few noise-canceling microphones are throat microphones.

## Connectors



Electronic symbol for a microphone

The most common connectors used by microphones are:

- Male XLR connector on professional microphones
- ¼ inch (sometimes referred to as 6.5 mm) jack plug also known as 1/4 inch TRS connector on less expensive consumer microphones. Many consumer microphones use an unbalanced 1/4 inch phone jack. Harmonica microphones commonly use a high impedance 1/4 inch TS connection to be run through guitar amplifiers.
- 3.5 mm (sometimes referred to as 1/8 inch mini) stereo (wired as mono) mini phone plug on very inexpensive and computer microphones

Some microphones use other connectors, such as a 5-pin XLR, or mini XLR for connection to portable equipment. Some lavalier (or 'lapel', from the days of attaching the microphone to the news reporters suit lapel) microphones use a proprietary connector for connection to a wireless transmitter. Since 2005, professional-quality microphones with USB connections have begun to appear, designed for direct recording into computer-based software.

## Impedance-matching

Microphones have an electrical characteristic called impedance, measured in ohms (Ω), that depends on the design. Typically, the *rated impedance* is stated. Low impedance is considered under 600 Ω. Medium impedance is considered between 600 Ω and 10 kΩ. High impedance is above 10 kΩ. Condenser microphones (after the built-in preamp) typically have an output impedance between 50 and 200 ohms.

The output of a given microphone delivers the same power whether it is low or high impedance. If a microphone is made in high and low impedance versions, the high impedance version has a higher output voltage for a given sound pressure input, and is suitable for use with vacuum-tube guitar amplifiers, for instance, which have a high input impedance and require a relatively high signal input voltage to overcome the tubes' inherent noise. Most professional microphones are low impedance, about 200 Ω or lower. Professional vacuum-tube sound equipment incorporates a transformer that steps up the impedance of the microphone circuit to the high impedance and voltage needed to drive the input tube; the impedance conversion inherently creates voltage gain as well. External matching transformers are also available that can be used in-line between a low impedance microphone and a high impedance input.

Low-impedance microphones are preferred over high impedance for two reasons: one is that using a high-impedance microphone with a long cable results in high frequency signal loss due to cable capacitance, which forms a low-pass filter with the microphone output impedance. The other is that long high-impedance cables tend to pick up more hum (and possibly radio-frequency interference (RFI) as well). Nothing is damaged if the impedance between microphone and other equipment is mismatched; the worst that happens is a reduction in signal or change in frequency response.

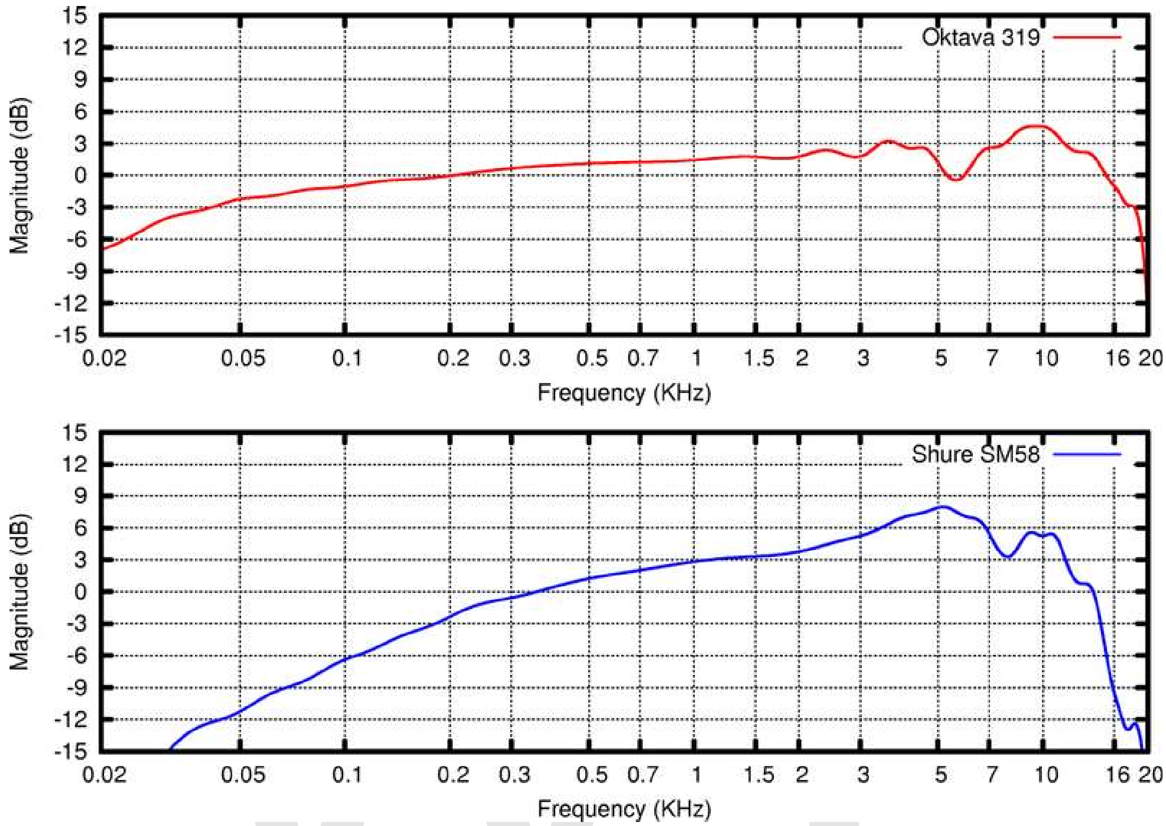
Most microphones are designed *not* to have their impedance matched by the load they are connected to. Doing so can alter their frequency response and cause distortion, especially at high sound pressure levels. Certain ribbon and dynamic microphones are exceptions, due to the designers' assumption of a certain load impedance being part of the internal electro-acoustical damping circuit of the microphone.

### Digital microphone interface

Neumann D-01 digital microphone and Neumann DMI-8 8-channel USB Digital Microphone Interface

The AES 42 standard, published by the Audio Engineering Society, defines a digital interface for microphones. Microphones conforming to this standard directly output a digital audio stream through an XLR male connector, rather than producing an analog output. Digital microphones may be used either with new equipment with appropriate input connections that conform to the AES 42 standard, or else via a suitable interface box. Studio-quality microphones that operate in accordance with the AES 42 standard are now available from a number of microphone manufacturers.

# Measurements and specifications




A comparison of the far field on-axis frequency response of the Oktava 319 and the Shure SM58

Because of differences in their construction, microphones have their own characteristic responses to sound. This difference in response produces non-uniform phase and frequency responses. In addition, microphones are not uniformly sensitive to sound pressure, and can accept differing levels without distorting. Although for scientific applications microphones with a more uniform response are desirable, this is often not the case for music recording, as the non-uniform response of a microphone can produce a desirable coloration of the sound. There is an international standard for microphone specifications, but few manufacturers adhere to it. As a result, comparison of published data from different manufacturers is difficult because different measurement techniques are used. The Microphone Data Website has collated the technical specifications complete with pictures, response curves and technical data from the microphone manufacturers for every currently listed microphone, and even a few obsolete models, and shows the data for them all in one common format for ease of comparison.. Caution should be used in drawing any solid conclusions from this or any other published data, however, unless it is known that the manufacturer has supplied specifications in accordance with IEC 60268-4.

A frequency response diagram plots the microphone sensitivity in decibels over a range of frequencies (typically at least 0–20 kHz), generally for perfectly on-axis sound (sound arriving at 0° to the capsule). Frequency response may be less informatively stated textually like so: "30 Hz–16 kHz ±3 dB". This is interpreted as meaning a nearly flat, linear, plot between the stated frequencies, with variations in amplitude of no more than plus or minus 3 dB. However, one cannot determine from this information how *smooth* the variations are, nor in what parts of the spectrum they occur. Note that commonly made statements such as "20 Hz–20 kHz" are meaningless without a decibel measure of tolerance. Directional microphones' frequency response varies greatly with distance from the sound source, and with the geometry of the sound source. IEC 60268-4 specifies that frequency response should be measured in *plane progressive wave* conditions (very far away from the source) but this is seldom practical. *Close talking* microphones may be measured with different sound sources and distances, but there is no standard and therefore no way to compare data from different models unless the measurement technique is described.

The self-noise or equivalent noise level is the sound level that creates the same output voltage as the microphone does in the absence of sound. This represents the lowest point of the microphone's dynamic range, and is particularly important should you wish to record sounds that are quiet. The measure is often stated in dB(A), which is the equivalent loudness of the noise on a decibel scale frequency-weighted for how the ear hears, for example: "15 dBA SPL" (SPL means sound pressure level relative to 20 micropascals). The lower the number the better. Some microphone manufacturers state the noise level using ITU-R 468 noise weighting, which more accurately represents the way we hear noise, but gives a figure some 11–14 dB higher. A quiet microphone typically measures 20 dBA SPL or 32 dB SPL 468-weighted. Very quiet microphones have existed for years for special applications, such the Brüel & Kjaer 4179, with a noise level around 0 dB SPL. Recently some microphones with low noise specifications have been introduced in the studio/entertainment market, such as models from Neumann and Røde that advertise noise levels between 5–7 dBA. Typically this is achieved by altering the frequency response of the capsule and electronics to result in lower noise within the A-weighting curve while broadband noise may be increased.

The maximum SPL (sound pressure level) the microphone can accept is measured for particular values of total harmonic distortion (THD), typically 0.5%. This amount of distortion is generally inaudible, so one can safely use the microphone at this SPL without harming the recording. Example: "142 dB SPL peak (at 0.5% THD)". The higher the value, the better, although microphones with a very high maximum SPL also have a higher self-noise.

The clipping level is perhaps a better indicator of maximum usable level, as the 1% THD figure usually quoted under max SPL is really a very mild level of distortion, quite inaudible especially on brief high peaks. Harmonic distortion from microphones is usually of low-order (mostly third harmonic) type, and hence not very audible even at 3–5%. Clipping, on the other hand, usually caused by the diaphragm reaching its absolute displacement limit (or by the preamplifier), produces a harsh sound on peaks, and should be avoided if at all possible. For some microphones the clipping level may be much higher than the max SPL.

The dynamic range of a microphone is the difference in SPL between the noise floor and the maximum SPL. If stated on its own, for example "120 dB", it conveys significantly less information than having the self-noise and maximum SPL figures individually.

Sensitivity indicates how well the microphone converts acoustic pressure to output voltage. A high sensitivity microphone creates more voltage and so needs less amplification at the mixer or recording device. This is a practical concern but is not directly an indication of the mic's quality, and in fact the term sensitivity is something of a misnomer, 'transduction gain' being perhaps more meaningful, (or just "output level") because true sensitivity is generally set by the noise floor, and too much "sensitivity" in terms of output level compromises the clipping level. There are two common measures. The (preferred) international standard is made in millivolts per pascal at 1 kHz. A higher value indicates greater sensitivity. The older American method is referred to a 1 V/Pa standard and measured in plain decibels, resulting in a negative value. Again, a higher value indicates greater sensitivity, so −60 dB is more sensitive than −70 dB.

# Measurement microphones

Some microphones are intended for testing speakers, measuring noise levels and otherwise quantifying an acoustic experience. These are calibrated transducers and are usually supplied with a calibration certificate that states absolute sensitivity against frequency. The quality of measurement microphones is often referred to using the designations "Class 1," "Type 2" etc., which are references not to microphone specifications but to sound level meters. A more comprehensive standard for the description of measurement microphone performance was recently adopted.

Measurement microphones are generally scalar sensors of pressure; they exhibit an omnidirectional response, limited only by the scattering profile of their physical dimensions. Sound intensity or sound power measurements require pressure-gradient measurements, which are typically made using arrays of at least two microphones, or with hot-wire anemometers.

### Microphone calibration

To take a scientific measurement with a microphone, its precise sensitivity must be known (in volts per pascal). Since this may change over the lifetime of the device, it is necessary to regularly calibrate measurement microphones. This service is offered by some microphone manufacturers and by independent certified testing labs. All microphone calibration is ultimately traceable to primary standards at a national measurement institute such as NPL in the UK, PTB in Germany and NIST in the USA, which most commonly calibrate using the reciprocity primary standard. Measurement microphones calibrated using this method can then be used to calibrate other microphones using comparison calibration techniques.

Depending on the application, measurement microphones must be tested periodically (every year or several months, typically) and after any potentially damaging event, such as being dropped (most such mikes come in foam-padded cases to reduce this risk) or exposed to sounds beyond the acceptable level.

## Microphone array and array microphones

A microphone array is any number of microphones operating in tandem. There are many applications:

- Systems for extracting voice input from ambient noise (notably telephones, speech recognition systems, hearing aids)
- Surround sound and related technologies
- Locating objects by sound: acoustic source localization, e.g. military use to locate the source(s) of artillery fire. Aircraft location and tracking.
- High fidelity original recordings
- 3D spatial beamforming for localized acoustic detection of subcutaneous sounds

Typically, an array is made up of omnidirectional microphones distributed about the perimeter of a space, linked to a computer that records and interprets the results into a coherent form.

## Microphone windscreens

Windscreens are used to protect microphones that would otherwise be buffeted by wind or vocal plosives from consonants such as "P", "B", etc. Most microphones have an integral windscreen built around the microphone diaphragm. A screen of plastic, wire mesh or a metal cage is held at a distance from the microphone diaphragm, to shield it. This cage provides a first line of defense against the mechanical impact of objects or wind. Some microphones, such as the Shure SM58, may have an additional layer of foam inside the cage to further enhance the protective properties of the shield. One disadvantage of all windscreen types is that the microphone's high frequency response is attenuated by a small amount, depending on the density of the protective layer.

Beyond integral microphone windscreens, there are three broad classes of additional wind protection.

### Microphone covers


Various microphone covers

Microphone covers are often made of soft open-cell polyester or polyurethane foam because of the inexpensive, disposable nature of the foam. Optional windscreens are often available from the manufacturer and third parties. A visible example of an optional accessory windscreen is the A2WS from Shure, one of which is fitted over each of the two Shure SM57 microphones used on the United States president's lectern. One disadvantage of polyurethane foam microphone covers is that they can deteriorate over time. Windscreens also tend to collect dirt and moisture in their open cells and must be cleaned to prevent high frequency loss, bad odor and unhealthy conditions for the person using the microphone. On the other hand, a major advantage of concert vocalist windscreens is that one can quickly change to a clean windscreen between users, reducing the chance of transferring germs. Windscreens of various colors can be used to distinguish one microphone from another on a busy, active stage.

## Pop filters

Pop filters or pop screens are used in controlled studio environments to minimize plosives when recording. A typical pop filter is composed of one or more layers of acoustically transparent gauze-like material, such as woven nylon (e.g. pantyhose) stretched over a circular frame and a clamp and a flexible mounting bracket to attach to the microphone stand. The pop shield is placed between the vocalist and the microphone. The need for a pop filter increases the closer a vocalist brings his or her lips to the microphone. Singers can be trained either to soften their plosives or direct the air blast away from the microphone, in which cases they don't need a pop filter.

Pop filters also keep spittle off the microphone. Most condenser microphones can be damaged by spittle.

## Blimps

Two recordings being made—A *blimp* is being used on the left. An open-cell foam windscreen is being used on the right.

Blimps (also known as Zeppelins) are large, hollow windscreens used to surround microphones for outdoor location audio, such as nature recording, electronic news gathering, and for film and video shoots. They can cut wind noise by as much as 25 dB, especially low-frequency noise. The blimp is essentially a hollow cage or basket with acoustically transparent material stretched over the outer frame. The blimp works by creating a volume of still air around the microphone. The microphone is often further isolated from the blimp by an elastic suspension inside the basket. This reduces wind vibrations and handling noise transmitted from the cage. To extend the range of wind speed conditions in which the blimp remains effective, many have the option of a secondary cover over the outer shell. This is usually an acoustically transparent, synthetic fur material with long, soft hairs. Common and slang names for this include "dead cat" or "windmuff". The hairs deaden the noise caused by the shock of wind hitting the blimp. A synthetic fur cover can reduce wind noise by an additional 10 dB.

**Chapter- 7**

# Air Core Gauge and Amplified Piezoelectric Actuator

## Air core gauge


An auto tachometer has a sweep of about 240-250 degrees and typically uses an air core gauge.

An **air core gauge** is a specific type of rotary actuator in an analog display gauge that allows an indicator to rotate a full 360 degrees. It is used in gauges and displays, most commonly automotive instrument clusters.

A typical automotive application is shown at the right. The air core gauge is a type of "air-core motor". It may be considered a "gauge movement" or "pointer indication device".

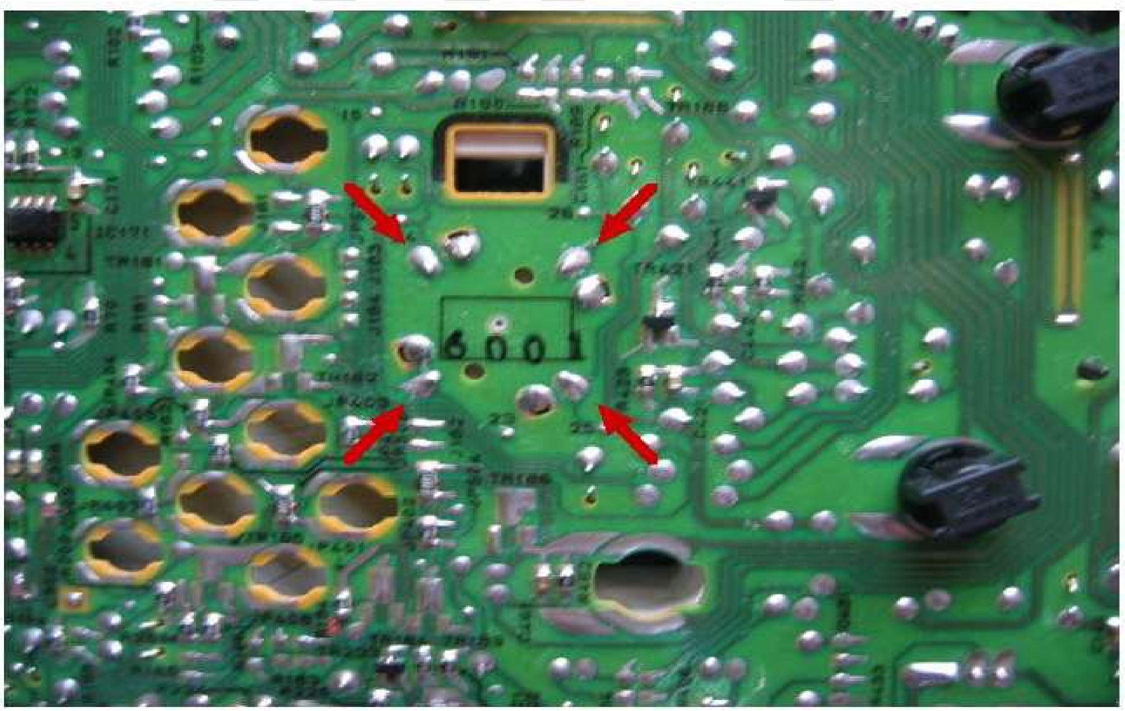
## Background

There are four common types of rotary actuators :

- Physical gauges, in which the needle is attached directly to the value being measured; for example, a mechanical pressure gauge
- Analog volt meters or d'Arsonval movements, which consist of a coil and a permanent magnet
- Stepper motors, which move in one-notch increments or steps
- Air-core motors, as described below.

## Construction and operation

The air core gauge consists of two independent, perpendicular coils surrounding a hollow chamber. A needle shaft protrudes into the chamber, where a permanent magnet is affixed to the shaft. When current flows through the perpendicular coils, their magnetic fields superimpose and the magnet is free to align with the combined fields.




Back side of an auto instrument cluster showing four mounting terminals for an air core gauge.

A typical air core gauge has four terminals, two for each coil, as shown. The two coils are identified as the sine coil and the cosine coil.

### Theory

The direction θ of the overall magnetic field is approximately:

$$\theta = \arctan\left(\frac{y}{x}\right)$$

Where *x* and *y* are the coils' respective currents. The permanent magnet aligns itself with that field, eventually settling near θ. In this way, by proportioning the current through each coil, the needle can reach all 360° of rotation.

### Example

If the sin coil current is 50 mA and the cos current is 29 mA:

The coil current ratio is 0.58, and arctan 0.58 = 30 degrees.

## Drivers

Air core gauges require special electronics to properly drive the coils. Driver integrated circuits typically have a serial input data port and two pair of output lines. One pair of the output lines drives the sin coil and one pair drives the cos coil.

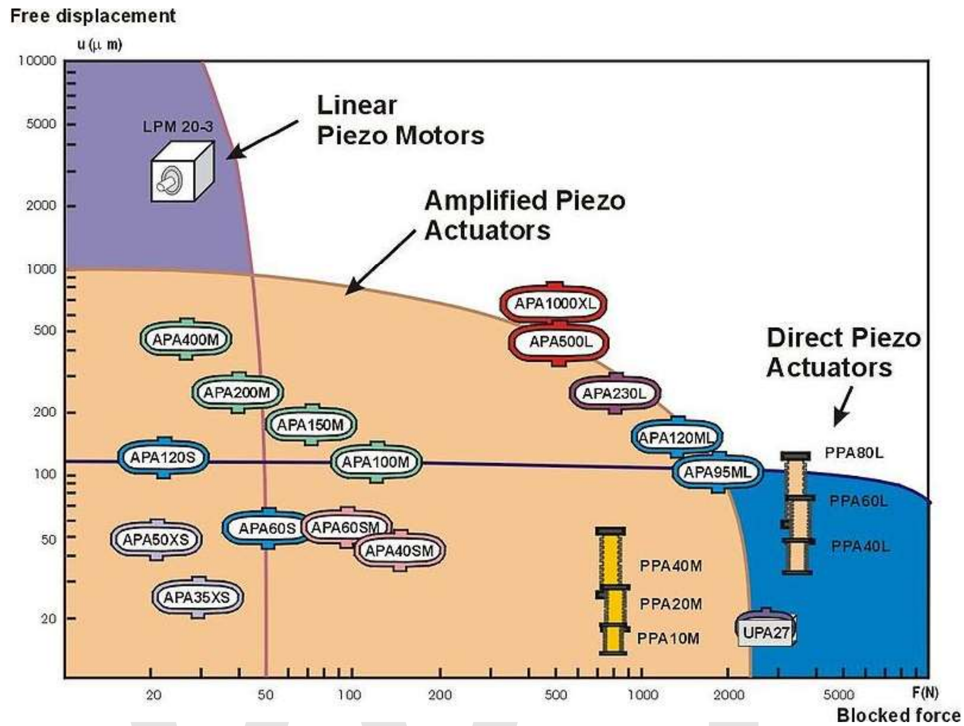
The input data defines:

- The quadrant to which the actuator will point. This defines the polarity of the voltage to the sin coil and the cos coil.
- The desired number of degrees within the quadrant.

Some typical driver ICs include:

- On Semiconductor CS4172 16 pin dual inline package
- On Semiconductor CS4192 surface mount package
- Melexis MLX10407
- Melexis MLX10420

# Amplified piezoelectric actuator




Panorama of piezoelectric actuators, classified on a stroke over force map

**Amplified piezoelectric actuators (APA)** are specific actuators using piezoelectric materials as active material, and having a specific design to overcome traditional limitations of classical direct piezoelectric actuator, the limited stroke. As classical piezoelectric materials have a strain of 0.1%, it is practically impossible to reach significant stroke without displacement amplification (1 mm displacement would required 1 meter of piezoelectric material). The solution to reach middle range stroke is to use an amplification system.

# Single-cell actuator

APA200M

APA$^{TM}$ is a trade mark of Cedrat Technologies using a patented design to create an amplification of the displacement. The principle is based on the deformation of an elliptic shell to amplify the ceramic strain. The ceramic stack is aligned with the great axis of the ellipse. A small deformation of the great axis creates a large displacement of the small axis. The amplification ratio can typically reach 20 times, that means such actuators can reach strokes of 1 mm.

The goal of the elliptic shell is not only to amplify the displacement. It has also to apply the correct pre-stress to the piezoelectric material in order to allow dynamic and precise motion. The other advantage is that this kind of flextentional actuators are very reliable.

## Multi-cell actuator

In diamond shaped amplifiers, using 4 piezo crystals instad of one increases control of movement, particularly in changing temperatures. More movers result in more force at similar displacement.

## Example of application

Piezoelectric actuators, and especially amplified piezo electric actuators are historically studied and used in aerospace application. NASA for example studied flextensional and test its own actuators for cryogenic application. Other organisations like ESA or ISRO are also studying such solution. Space agency's interest for amplified piezoelectric actuators is due to the high power density of these actuators, their reliability and low power losses when used in quasi-static operation. This is definitely a key advantage for vacuum application. Compare for example to magnetic actuators. Controlling helicopter rotor blades using active flaps has been investigated for some time without being put into production, and *amplified piezoelectric actuator*s is the most common technology used.

## Other solutions

Other solutions are possible to amplify the stroke within the actuator. The first solution is to use lever arm. This classical solution is simple in its principle, but requires specific know-how to be reliable. Its main disadvantage is it can hardly allow dynamic displacement. One other solution is to use a specific shell for the displacement amplification, typically known as a dog bone, and to had an external pre-stress. As the pre-stress is not controlled by the shell, it has to have a stifness as low as possible. One classical solution is to use flexural hinge to reduce the shell stiffness.

## Chapter- 8

# Ball Screw and Carbon Nanotube Actuators

## Ball screw

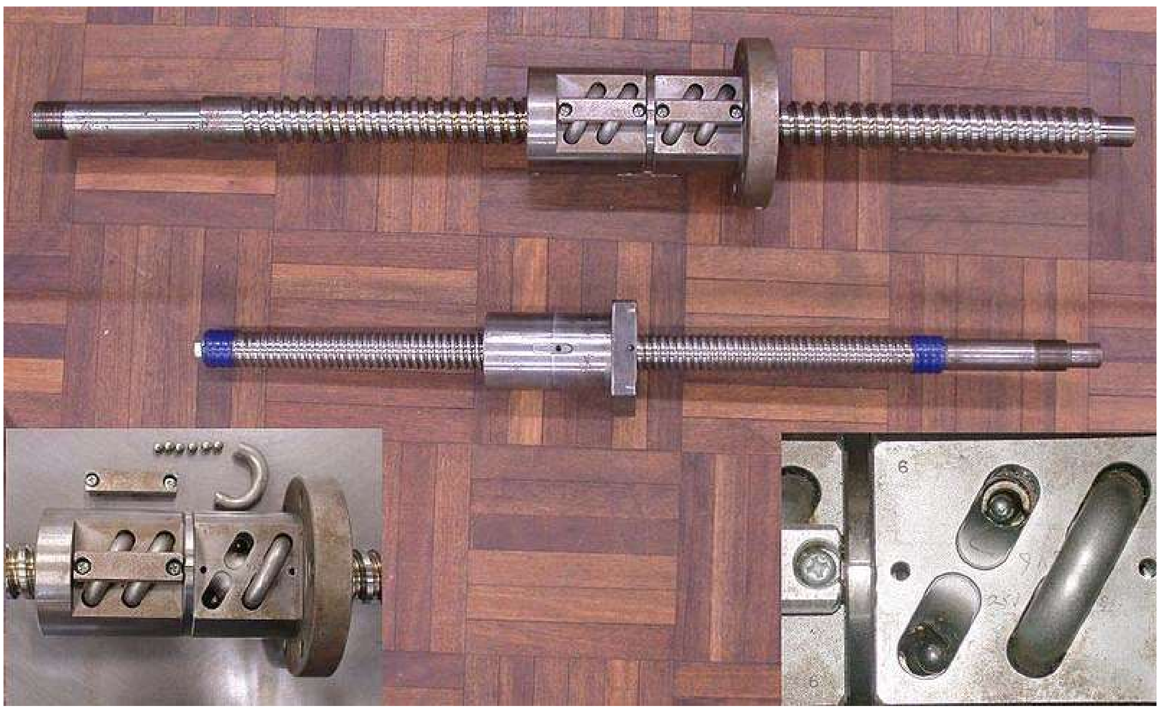


Photo showing two ball screws. Inset images are close-up photos of the ball assembly of the top screw. Left inset: recirculating tube removed showing retainer bracket, loose balls and tube. Right inset: closer view of the nut cavity.

A **ball screw** is a mechanical linear actuator that translates rotational motion to linear motion with little friction. A threaded shaft provides a helical raceway for ball bearings which act as a precision screw. As well as being able to apply or withstand high thrust loads, they can do so with minimum internal friction. They are made to close tolerances and are therefore suitable for use in situations in which high precision is necessary. The ball assembly acts as the nut while the threaded shaft is the screw.

Ball screws are used in aircraft and missiles to move control surfaces, especially for electric fly by wire. They are also used in machine tools, robots and precision assembly equipment. High precision ball screws are used in steppers for semiconductor manufacturing.

In contrast to conventional leadscrews, ballscrews tend to be rather bulky, due to the need to have a mechanism to re-circulate the balls.

To maintain their inherent accuracy and ensure long life, great care is needed to avoid contamination with dirt and abrasive particles. This may be achieved by using rubber or leather bellows to completely or partially enclose the working surfaces. Another solution is to use a positive pressure of filtered air when they are used in a semi-sealed or open enclosure.

While reducing friction, ball screws can operate with some preload, effectively eliminating backlash (slop) between input (rotation) and output (linear motion). This feature is essential when they are used in computer-controlled motion-control systems, e.g. CNC machine tools and high precision motion applications (e.g. wire bonding).

Depending upon their lead angle, ball screws can be back-driven due to their low internal friction (i.e. the screw shaft can be driven linearly to rotate the ball nut). They are usually undesirable for hand-fed machine tools, as the stiffness of a servo motor is required to keep the cutter from grabbing the work and self feeding, that is, where the cutter and workpiece exceed the optimum feedrate and effectively jam or crash together, ruining the cutter and workpiece. Cost is also a major factor as Acme screws are cheaper to manufacture.

Low friction in ball screws yields high mechanical efficiency compared to alternatives. A typical ball screw may be 90 percent efficient, versus 50 percent efficiency of an Acme lead screw of equal size. The higher cost of ball screws may thus be offset by lower power requirements for the same net performance.

Ball screw shafts may be fabricated by rolling, yielding a less precise, but inexpensive and mechanically efficient product. Rolled ball screws have a positional precision of several thousandths of an inch per foot.

High-precision screw shafts are typically precise to one thousandth of an inch per foot or better. They have historically been machined to gross shape, case hardened and then ground. The three step process is needed because high temperature machining distorts the work-piece. Hard whirling is a recent (2008) precision machining technique that minimizes heating of the work, and can produce precision screws from case-hardened bar stock.

Instrument quality screw shafts are typically precise to 250 nanometers per centimeter. They are produced on precision milling machines with optical distance measuring equipment and special tooling. Similar machines are used to produce optical lenses and mirrors. Instrument screw shafts are generally made of Invar, to prevent temperature from changing tolerances too much.

Historically, the first precise screwshafts were produced by starting with a low precision screwshaft, and then lapping the shaft with several spring-loaded nut laps. By rearranging and inverting the nut laps, the lengthwise errors of the nuts and shaft were averaged. Then, the very repeatable shaft's pitch is measured against a distance standard. A similar process is sometimes used today to produce reference standard screw shafts, or master manufacturing screw shafts.

## Ball return systems

The bearing balls travel inside the screw and nut thread. If the ball nut did not have a return mechanism the balls would fall out of the end of the ball nut when they reached the end of the nut. For this reason several different recirculation methods have been developed.

An external ballnut employs a stamped tube which picks up balls from raceway with use of small pick up finger. Balls travel inside of tube and are then replaced back in thread raceway.

An internal button ballnut employs a machined or cast button style return which allows balls to exit raceway track and move one thread and reenter raceway.

An endcap return ballnut employs a cap on the end of ball nut. The cap is machined to pick up balls out of the end of nut and direct them down holes which are bored transversely down the ballnut. The compliment cap on the other side of nut directs balls back into raceway.

## Equations

$$T = \frac{Fl}{2\pi\nu}$$

Where $T$ is torque applied to screw or nut, $F$ is linear force applied, $l$ is ball screw lead, and $\nu$ is ball screw efficiency.

# Carbon nanotube actuators

The exceptional electrical and mechanical properties of **carbon nanotubes** have made them alternatives to the traditional electrical actuators for both microscopic and macroscopic applications. Carbon nanotubes are very good conductors of both electricity and heat, and they are also very strong and elastic molecules in certain directions. These properties are difficult to find in the same material and very needed for high performance actuators. For current carbon nanotube actuators, multi-walled carbon nanotubes (MWNTs) and bundles of MWNTs have been widely used mostly due to the easiness of handling and robustness. Solution dispersed thick films and highly ordered transparent films of carbon nanotubes have been used for the macroscopic applications.

## Microscopic applications

### Carbon nano-tweezers

Carbon nanotube tweezers have been fabricated by deposition of MWNT bundles on isolated electrodes deposited on tempered glass micropipettes Those nanotube budles can be mechanically manipulated by electricity and can be used to manipulate and transfer micro- and nano-structures. The nanotube bundles used for tweezers are about 50 nm in diameter and 2 μm in lengths. Under electric bias, two close sets of bundles are attracted and can be used as nanoscale tweezers.

### Nanotube on/off switches and random access memory

Harvard researchers have used the electrostatic attraction principle to design on/off switches for their proposed nanotube Random Access Memory devices. They used carbon nanotube bundles of ~50 nm in diameter to fabricate their proof-of-concept prototypes. One set of MWNT bundles are laid on the substrate and another set of bundles is trenched on top of the underlying nanotube bundles with air gap in between. Once electrical bias is applied the sets of nanotube bundles are attracted, thus changing the electrical resistance. These two states of resistance are on and off states. They have managed to get more than 10 times difference between off and on state resistances. This idea can be used as very highly packed arrays of nanoswitches and random access memory deivces if they can be applied to arrays of single-walled carbon nanotubes, which are about 1 nm in diameter and hundreds of micrometres in length. The current technical challenge with this design is the lack of control to place arrays of carbon nanotubes on substrate.

# Macroscopic applications

## Nanotube sheet electrodes as actuators

Researchers of AlliedSignal initially demonstrated the possibility of electrically powered actuators fabricated by carbon nanotube sheets. They taped carbon nanotube sheets on two sides of a double sided scotch tape and applied potential on the nanotube sheets in a NaCl electrolyte solution. Nanotube sheets are used as electrolyte-filled electrodes of a super capacitor. Nanotube sheets are electrically charged by the double layer formation at the nanotube-electrolyte interface without any need of ion intercalation. Therefore electrically driven actuators of nanotube sheets are superior to the conjugated polymer actuators which involve solid-state dopant diffusion and structural changes limiting rate, cycle life and energy conversion efficiencies. On the other hand, ferroelectric and electrostrictive materials are also very useful for direct energy conversion, but they require high operation voltages and ambient temperature of a limited range. Nanotube sheet actuators were shown to operate at low voltages (~1 Volts or less) and provide higher work densities per cycle than other alternative technologies. Later Baughman et al. showed that actuator response can be observed up to switching rates of 1 kHz and cycling the nanotube actuator at constant rate of 1 Hz for 140000 cycles decreases the stroke by ~33%. 0.75 MPa of stress were measured on the nanotube sheet actuators, which is greater than the maximum stress (0.3 MPa) that can be loaded on a human muscle.

The maximum actuator strain for electrically driven actuators of carbon nanotube sheets can be improved up to 0.7% in a 1 M electrolyte once the sheets are annealed in an inert atmosphere at very high temperatures (1100 $^{o}$C) in contrast to once reported 0.1% or less for low electrochemical potentials (~1 V or less). The maximum strain for the carbon nanotube sheet actuators at low voltages is greater than that of the high modulus ferroelectric ceramic actuators (~0.1%), but it is lower than that of the low voltage (~0.4 V) conducting polymer actuators (~3% film direction, 20% thickness direction). Strokes were reported as high as 215% for strain biased low modulus electrostrictive rubbers under biases greater than 1kV (corresponding to an electric field 239 MV/m for the geometry mentioned in the reference paper). Spinks et al. realized pneumatic actuation from the carbon nanotube sheets in electrolyte solutions with high electrochemical potential (1.5 V), which cause gas generation in the electrolyte. The released gas dramatically increases the actuator stroke from the carbon nanotube sheet. Thickness of the carbon nanotube sheet expands by ~300% and the sheet plane contracts by 3%.

## Artificial muscles and giant strokes by MWNT aerogel sheets

Highly ordered free standing aerogel sheets of MWNTs can be realized by simply drawing the sheet from the sidewalls of CVD grown MWNT forests. UT researchers came up with the conventional method where they attach an adhesive tape to the sidewalls of MWNT forests and they pull the tape at a constant rate as fast as 7 meters per minute to get 3–5 cm wide aerogel sheets of aligned MWNTs which have exceptional mechanical and optical properties. The aerogel sheets have a density of ~1.5 mg/cm$^3$, an areal density of 1-3 μg/cm$^2$ and a thickness of ~20 μm. The thickness is decreased to

~50 nm by liquid-based densification to decrease the volume. The aerogel sheets can be stretched as much as three times along the width while low-modulus rubber like behavior is remained.

Having aerogel sheets of MWNTs, UT researchers fabricated actuators with giant strokes (~180% actuation along the width) with 5 ms delay time between applying the potential and observing the maximum stroke. Therefore the actuation rate is slightly better than that of the human muscle. This is a very important achievement considering the actuation rate for artificial muscles used in robots is typically much slower. Furthermore the use of carbon nanotubes as the building blocks as an artificial muscle also helps in terms of strength and robustness by making the artificial muscle stronger than steel in one direction and more flexible than rubber in the other two directions. The lack of electrolyte solution and temperature robustness of the aerogel sheet in inert ambient makes high temperature operation possible. The actuation stroke decreases by only 50% from its room temperature value to 1344 $^{o}$C. Thus, this design of artificial muscles can be quite useful for many industrial applications with the drawback of high voltage operation for giant strokes.

## Challenges and future applications

As a result, carbon nanotubes have been shown to be great materials for actuation related applications. The subfield of carbon nanotube actuators have been quite successful and ready for scalable applications considering there are quite a few conventional and scalable methods for the synthesis of large scale carbon nanotubes. Carbon nanotube sheets used as electrodes in electrolyte solutions offered low voltage operations at room temperature with actuation strokes and rates comparable to the conducting polymer actuators, but with higher work densities per cycle and life times. However the actuation strokes are much smaller than those of the electrostrictive rubbers which operate at three orders of magnitude higher voltages. On the other hand, realization of carbon nanotube aerogels made giant strokes possible comparable to electrostrictive rubbers at room temperature, but carbon nanotube aerogels can perform at a very wide range of temperatures, and with very high actuation rates, which are even better than the actuation rate of the human muscles.

## Chapter- 9

# Hoist (Device)


Overhead crane with a Wire Rope **hoist (device)** in blue suspended from bridge crane girder

A **hoist** is a device used for lifting or lowering a load by means of a drum or lift-wheel around which rope or chain wraps. It may be manually operated, electrically or pneumatically driven and may use chain, fiber or wire rope as its lifting medium. The load is attached to the hoist by means of a lifting hook.

## Types of Hoist

The basic hoist has two important characteristics to define it: Lifting medium and power type. The lifting medium is either wire rope, wrapped around a drum, or load-chain, raised by a pulley with a special profile to engage the chain. The power type can be either electric motor or air motor. Both the wire rope hoist and chain hoist have been in common use since the 1800s. however; Mass production of an electric hoist did not start until the early 1900's and was first adapted by Germany. A hoist can be built as one integral-package unit, designed for cost-effective purchasing and moderate use, or it can be built as a built-up custom unit, designed for durability and performance. The built-up hoist will be much more expensive, but will also be easier to repair and more durable. Package units are where once regarded as being designed for light to moderate usage, but since the 60's this has changed. Built-up units are designed for heavy to severe service, but over the years that market has decreased in size since the advent of the more durable packaged hoist. A machine shop or fabricating shop will use an integral-package hoist, while a Steel Mill or NASA would use a built-up unit to meet durability, performance, and repairability requirements. NASA has also seen a change in the use of package hoists. The NASA Astronaut training pool for example utilizes cranes with packaged hoists.

**Wire Rope Hoist or Chain Hoist**


Wire Rope hoist on an overhead crane being used in typical machine shop. The hoist is operated via a wired pushbutton station to move system and the load in any direction

Builder's hoist, with small petrol engine

More commonly used hoist in today's worldwide market is an electrically powered hoist. These are either the chain type or the wire rope type. **Demag Cranes & Components Corp.** was one of the first companies in the world to mass-produce hoists. The first units large in size date back in 1819 and was powered by steam.

Now many hoists are package hoists, built as one unit in a single housing, generally designed for ten-year life, but the life calculation is based on an industry standard when calculating actual life. In today's modern world for the North American market there are a few governing bodies for the industry. The Overhead Alliance is a group that represents Crane Manufacturers Association of America (CMAA), Hoist Manufacturers Institute (HMI), and Monorail Manufacturers Association (MMA). These product counsels of the

Material Handling Industry of America have joined forces to create promotional materials to raise the awareness of the benefits to overhead lifting. The members of this group are marketing representatives of the member companies.

Common small portable hoists are of two main types, the *chain hoist* or *chain block* and the wire rope or cable type. Chain hoists may have a lever to actuate the hoist or have a loop of operating chain that one pulls through the block (known traditionally as a chain fall) which then activates the block to take up the main lifting chain.

A hand powered hoist with a ratchet wheel is known as a "ratchet lever hoist" or, colloquially, a "Come-A-Long". The original hoist of this type was developed by Abraham Maasdam of Deep Creek, Colorado about 1919, and later commercialized by his son, Felber Maasdam, about 1946. It has been copied by many manufacturers in recent decades. A similar heavy duty unit with a combination chain and cable became available in 1935 that was used by railroads, but lacked the success of the cable only type units.

A ratchet lever hoist (Come-A-Long).

Ratchet lever hoists have the advantage that they can usually be operated in any orientation, for pulling, lifting or binding. Chain block type hoists are usually suitable only for vertical lifting.

For a given *rated load* wire rope is lighter in weight per unit length but overall length is limited by the drum diameter that the cable must be wound onto. The lift chain of a chain hoist is far larger than the liftwheel over which chain may function. Therefore, a high-performance chain hoist may be of significantly smaller physical size than a wire rope hoist rated at the same working load.

Both systems fail over time through fatigue fractures if operated repeatedly at loads more than a small percentage of their tensile breaking strength. Hoists are often designed with internal clutches to limit operating loads below this threshold. Within such limits wire rope rusts from the inside outward while chain links are markedly reduced in cross section through wear on the inner surfaces. Regular lubrication of both tensile systems is recommended to reduce frequency of replacement. High speed lifting, greater than about 60 feet per minute (18.3 m/min), requires wire rope wound on a drum, because chain over a pocket wheel generates fatigue-inducing resonance for long lifts.

The unloaded wire rope of small hand powered hoists often exhibits a snarled "set", making the use of a chain hoist in this application less frustrating, but heavier. In addition, if the wire in a wire hoist fails, it can whip and cause injury, while a chain will simply break.

## Construction hoists

A hoist on the Trump International Hotel & Tower-Chicago

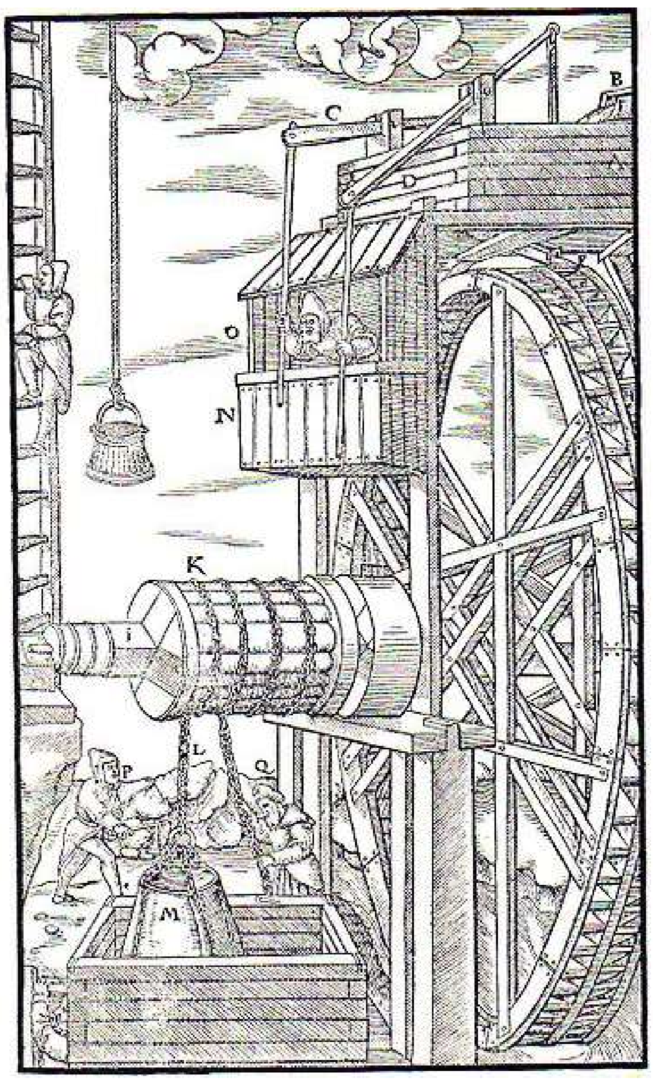
Also known as a Man-Lift, Buckhoist, temporary elevator, Alimak or construction elevator, this type of hoist is commonly used on large scale construction projects, such as high-rise buildings or major hospitals. There are many other uses for the construction elevator. Many other industries use the buckhoist for full time operations. The purpose being to carry personnel, materials, and equipment quickly between the ground and higher floors, or between floors in the middle of a structure.

The construction hoist is made up of either one or two cars (cages) which travel vertically along stacked mast tower sections. The mast sections are attached to the structure or building every 25 feet (7.62 m) for added stability. For precisely controlled travel along the mast sections, modern construction hoists use a motorized rack-and-pinion system that climbs the mast sections at various speeds.

While hoists have been predominantly produced in Europe and the United States, China is emerging as a manufacturer of hoists to be used in Asia.

In the United States and abroad, General Contractors and various other industrial markets rent or lease hoists for a specific projects. Rental or leasing companies provide erection, dismantling, and repair services to their hoists to provide General Contractors with turnkey services. Also the rental and leasing companies can provide parts and service for the elevators that are under contract.

## Mine hoists




A water-powered mine hoist used for raising ore from *De re metallica*.

In underground mining a **hoist** or **winder** is used to raise and lower conveyances within the mine shaft. Human, animal and water power were used to power the mine hoists documented in Agricola's De Re Metallica, published in 1556. Stationary steam engines were commonly used to power mine hoists through the 19th century and into the 20th, as at the Quincy Mine, where a 4-cylinder cross-compound corliss engine was used. Modern hoists are powered using electric motors, historically with direct current drives utilizing solid-state converters (thyristors), however modern large hoists use alternating current drives that are variable frequency controlled. There are three principal types of hoists used in mining applications, Drum Hoists, Friction (or Kope) hoists and Blair multi-rope hoists.

## Chapter- 10

# Hydraulic Cylinder


The hydraulic cylinders on this excavator control the machine's linkages.

A **Hydraulic cylinder** (also called a linear hydraulic motor) is a mechanical actuator that is used to give a unidirectional force through a unidirectional stroke. It has many applications, notably in engineering vehicles.

## Operation

Hydraulic cylinders get their power from pressurized hydraulic fluid, which is typically oil. The hydraulic cylinder consists of a cylinder barrel, in which a piston connected to a piston rod moves back and forth. The barrel is closed on each end by the cylinder bottom (also called the cap end) and by the cylinder head where the piston rod comes out of the cylinder. The piston has sliding rings and seals. The piston divides the inside of the cylinder in two chambers, the bottom chamber (cap end) and the piston rod side chamber (rod end). The hydraulic pressure acts on the piston to do linear work and motion.

Flanges, trunnions, and/or clevisses are mounted to the cylinder body. The piston rod also has mounting attachments to connect the cylinder to the object or machine component that it is pushing.

A hydraulic cylinder is the actuator or "motor" side of this system. The "generator" side of the hydraulic system is the hydraulic pump which brings in a fixed or regulated flow of oil to the bottom side of the hydraulic cylinder, to move the piston rod upwards. The piston pushes the oil in the other chamber back to the reservoir. If we assume that the oil pressure in the piston rod chamber is approximately zero, the force $F$ on the piston rod equals the pressure $P$ in the cylinder times the piston area $A$:

$$F = P \cdot A.$$

The piston moves instead downwards if oil is pumped into the piston rod side chamber and the oil from the piston area flows back to the reservoir without pressure. The pressure in the piston rod area chamber is (Pull Force) / (piston area - piston rod area).

# Parts of a hydraulic cylinder

A hydraulic cylinder consists of the following parts:

## Cylinder barrel

The cylinder barrel is mostly a seamless thick walled forged pipe that must be machined internally. The cylinder barrel is ground and/or honed internally.

## Cylinder Bottom or Cap

In most hydraulic cylinders, the barrel and the bottom portion are welded together. This can damage the inside of the barrel if done poorly. Therefore some cylinder designs have a screwed or flanged connection from the cylinder end cap to the barrel. In this type the barrel can be disassembled and repaired in future.

## Cylinder Head

The cylinder head is sometimes connected to the barrel with a sort of a simple lock (for simple cylinders). In general however the connection is screwed or flanged. Flange connections are the best, but also the most expensive. A flange has to be welded to the pipe before machining. The advantage is that the connection is bolted and always simple to remove. For larger cylinder sizes, the disconnection of a screw with a diameter of 300 to 600 mm is a huge problem as well as the alignment during mounting.

## Piston

The piston is a short, cylinder-shaped metal component that separates the two sides of the cylinder barrel internally. The piston is usually machined with grooves to fit elastomeric or metal seals. These seals are often O-rings, U-cups or cast iron rings. They prevent the pressurized hydraulic oil from passing by the piston to the chamber on the opposite side. This difference in pressure between the two sides of the piston causes the cylinder to extend and retract. Piston seals vary in design and material according to the pressure and temperature requirements that the cylinder will see in service. Generally speaking, elastomeric seals made from nitrile rubber or other materials are best in lower temperature environments while seals made of Viton are better for higher temperatures. The best seals for high temperature are cast iron piston rings.

## Piston Rod

The piston rod is typically a hard chrome-plated piece of cold-rolled steel which attaches to the piston and extends from the cylinder through the rod-end head. In double rod-end cylinders, the actuator has a rod extending from both sides of the piston and out both ends of the barrel. The piston rod connects the hydraulic actuator to the machine component doing the work. This connection can be in the form of a machine thread or a mounting attachment such as a rod-clevis or rod-eye. These mounting attachments can be threaded or welded to the piston rod or, in some cases, they are a machined part of the rod-end.

## Rod gland

The cylinder head is fitted with seals to prevent the pressurized oil from leaking past the interface between the rod and the head. This area is called the rod gland. It often has another seal called a rod wiper which prevents contaminants from entering the cylinder when the extended rod retracts back into the cylinder. The rod gland also has a rod wear ring. This wear ring acts as a linear bearing to support the weight of the piston rod and guides it as it passes back and forth through the rod gland. In some cases, especially in small hydraulic cylinders, the rod gland and the rod wear ring are made from a single integral machined part.

## Other parts

- Cylinder bottom connection
- Seals
- Cushions

A hydraulic cylinder should be used for pushing and pulling only. No bending moments or side loads should be transmitted to the piston rod or the cylinder. For this reason, the ideal connection of a hydraulic cylinder is a single clevis with a spherical ball bearing. This allows the hydraulic actuator to move and allow for any misalignment between the actuator and the load it is pushing.

# Hydraulic Cylinder Designs

There are primarily two styles of hydraulic cylinder construction used in industry: tie rod style cylinders and welded body style cylinders.

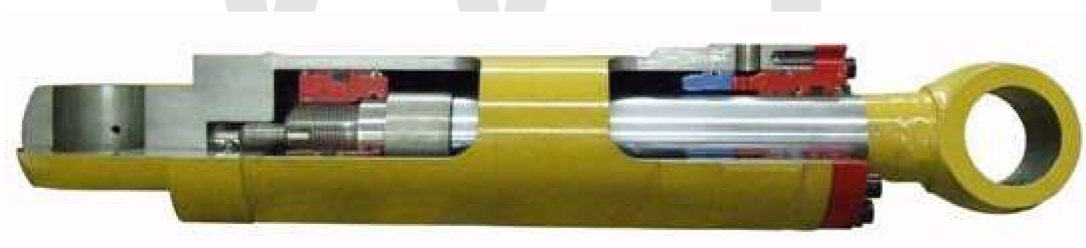
## Tie Rod Cylinders

Tie rod style hydraulic cylinders use high strength threaded steel rods to hold the two end caps to the cylinder barrel. This method of construction is most often seen in industrial factory applications. Small bore cylinders usually have 4 tie rods, while large bore cylinders may require as many as 16 or 20 tie rods in order to retain the end caps under the tremendous forces produced. Tie rod style cylinders can be completely disassembled for service and repair.

The National Fluid Power Association (NFPA) has standardized the dimensions of hydraulic tie rod cylinders. This enables cylinders from different manufacturers to interchange within the same mountings.

## Welded Body Cylinders

Welded body cylinders have no tie rods. The barrel is welded directly to the end caps. The ports are welded to the barrel. The front rod gland is usually threaded into or bolted to the cylinder barrel. This allows the piston rod assembly and the rod seals to be removed for service.



A Cut Away of a Welded Body Hydraulic Cylinder showing the internal components

Welded body cylinders have a number of advantages over tie rod style cylinders. Welded cylinders have a narrower body and often a shorter overall length enabling them to fit better into the tight confines of machinery. Welded cylinders do not suffer from failure due to tie rod stretch at high pressures and long strokes. The welded design also lends itself to customization. Special features are easily added to the cylinder body. These may include special ports, custom mounts, valve manifolds, and so on.

The smooth outer body of welded cylinders also enables the design of multi-stage telescopic cylinders.

Welded body hydraulic cylinders dominate the mobile hydraulic equipment market such as construction equipment (excavators, bulldozers, and road graders) and material handling equipment (forklift trucks, telehandlers, and lift-gates). They are also used in heavy industry such as cranes, oil rigs, and large off-road vehicles in above-ground mining.

# Piston Rod construction

The piston rod of a hydraulic cylinder operates both inside and outside the barrel, and consequently both in and out of the hydraulic fluid and surrounding atmosphere.

## Metallic coatings

Smooth and hard surfaces are desirable on the outer diameter of the piston rod and slide rings for proper sealing. Corrosion resistance is also advantageous. A chromium layer may often be applied on the outer surfaces of these parts. However, chromium layers may be porous, thereby attracting moisture and eventually causing oxidation. In harsh marine environments, the steel is often treated with both a nickel layer and a chromium layer. Often 40 to 150 micrometer thick layers are applied. Sometimes solid stainless steel rods are used. High quality stainless steel such as AISI 316 may be used for low stress applications. Other stainless steels such as AISI 431 may also be used where there are higher stresses, but lower corrosion concerns.

## Ceramic coatings

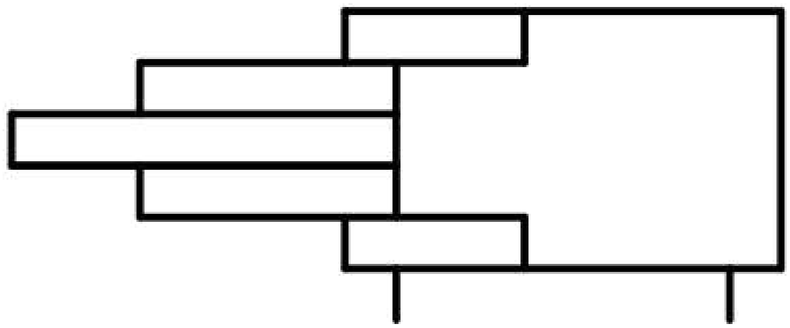
Due to shortcomings of metallic materials, ceramic coatings were developed. Initially ceramic protection schemes seemed ideal, but porosity was higher than projected. Recently the corrosion resistant semi ceramic Lunac 2+ coatings were introduced. These hard coatings are non porous and do not suffer from high brittleness.

## Lengths

Piston rods are generally available in lengths which are cut to suit the application. As the common rods have a soft or mild steel core, their ends can be welded or machined for a screw thread.

# Special hydraulic cylinders

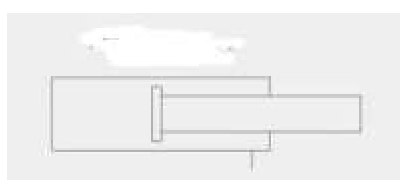
## Telescopic cylinder



Telescopic cylinder (ISO 1219 symbol)

The length of a hydraulic cylinder is the total of the stroke, the thickness of the piston, the thickness of bottom and head and the length of the connections. Often this length does not fit in the machine. In that case the piston rod is also used as a piston barrel and a second piston rod is used. These kind of cylinders are called telescopic cylinders. If we call a normal rod cylinder single stage, telescopic cylinders are multi-stage units of two, three, four, five and even six stages. In general telescopic cylinders are much more expensive than normal cylinders. Most telescopic cylinders are single acting (push). Double acting telescopic cylinders must be specially designed and manufactured.

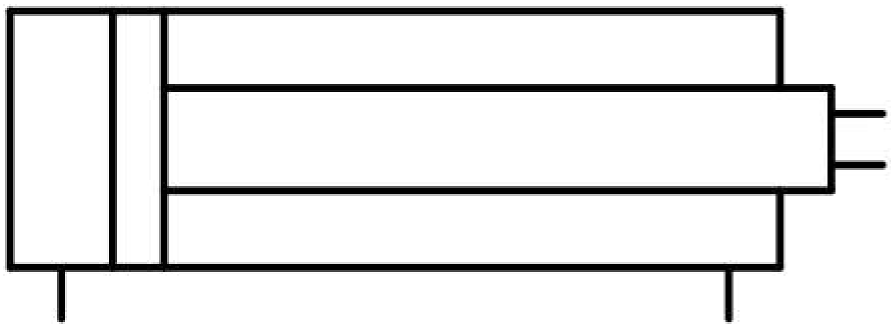
## Plunger cylinder



Plunger cylinder

A hydraulic cylinder without a piston or with a piston without seals is called a plunger cylinder. A plunger cylinder can only be used as a pushing cylinder; the maximum force is piston rod area multiplied by pressure. This means that a plunger cylinder in general has a relatively thick piston rod.

## Differential cylinder



Differential cylinder (ISO 1219 symbol)

A differential cylinder acts like a normal cylinder when pulling. If the cylinder however has to push, the oil from the piston rod side of the cylinder is not returned to the reservoir, but goes to the bottom side of the cylinder. In such a way, the cylinder goes much faster, but the maximum force the cylinder can give is like a plunger cylinder. A differential cylinder can be manufactured like a normal cylinder, and only a special control is added.

## Rephasing cylinder

Rephasing cylinders are two or more cylinders plumbed in series or parallel, with the bores and rods sized such that all rods extend and/or retract equally when flow is directed to the first, or last, cylinder within the system.

In "parallel" applications, the bore and rod sizes are always the same, and the cylinders are always used in pairs. In "series" applications, the bore and rod sizes are always different, and two or more cylinders may be used. In these applications, the bores and rods are sized such that all rods extend or retract equally when flow is applied to the first or last cylinder within the system.

This hydraulic synchronization of rod positions eliminates the need for a flow divider in the hydraulic system, or any type of mechanical connection between the cylinder rods to achieve synchronization.

## Position sensing "smart" hydraulic cylinder

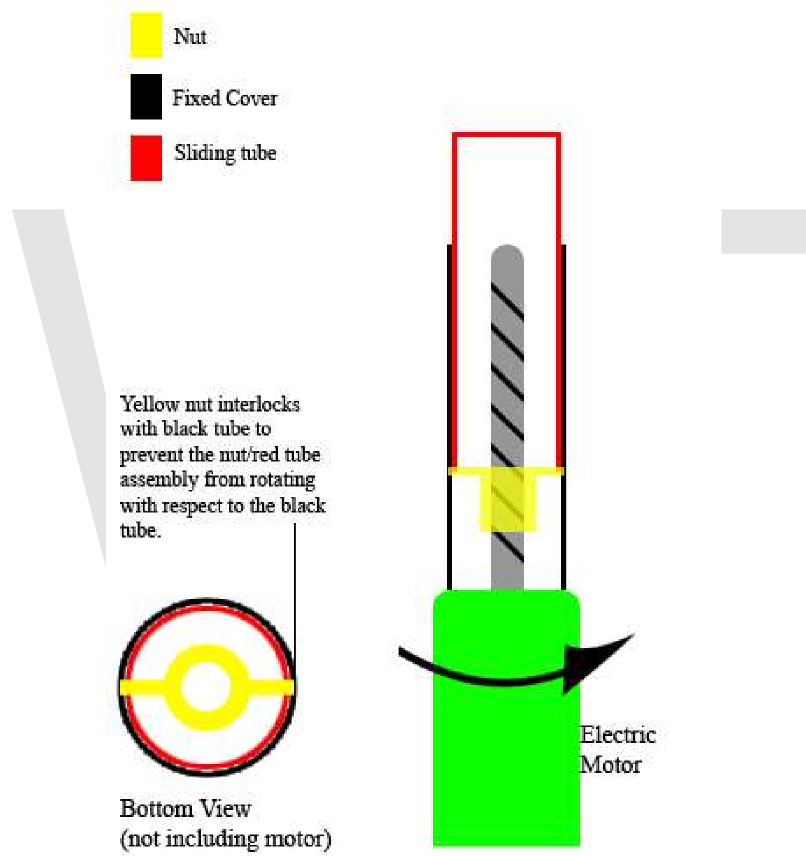
Position sensing hydraulic cylinders eliminate the need for a hollow cylinder rod. Instead, an external sensing “bar” utilizing Hall-Effect technology senses the position of the cylinder’s piston. This is accomplished by the placement of a permanent magnet within the piston. The magnet propagates a magnetic field through the steel wall of the cylinder, providing a locating signal to the sensor.

## A note about popular terminology

At least in the USA, popular usage sometimes refers to the whole assembly of cylinder, piston, and piston rod (or more) collectively as a "piston", which is incorrect. See, for instance, "Hydraulic piston raises the table from 19 (in.) to 26 (in.)" Marine Tables, Inc. (Select item 3 of 8, near the bottom.)

## Chapter- 11

# Linear Actuator




Conceptual design of a basic traveling-nut linear actuator. Note that in this example the lead screw (gray) rotates while the lead nut (yellow) and tube (red) do not.

A **linear actuator** is an actuator that, when driven by a non-linear motion, creates linear motion (as opposed to rotary motion, e.g. of an electric motor). Mechanical and hydraulic actuation are the most common methods of achieving the linear motion.

# Types

## Mechanical actuators


A mechanical linear actuator with digital readout.

Mechanical linear actuators operate by conversion of rotary motion into linear motion. Conversion is commonly made via a few simple types of mechanism:

- **Screw**: Screw jack, ball screw and roller screw actuators all operate on the principle of the simple machine known as the screw. By rotating the actuator's nut, the screw shaft moves in a line.
- **Wheel and axle**: Hoist, winch, rack and pinion, chain drive, belt drive, rigid chain and rigid belt actuators operate on the principle of the wheel and axle. By rotating a wheel/axle (e.g. drum, gear, pulley or shaft) a linear member (e.g. cable, rack, chain or belt) moves.
- **Cam**: Cam actuators function on a principle similar to that of the wedge, but provide relatively limited travel. As a wheel-like cam rotates, its eccentric shape provides thrust at the base of a shaft.

Some mechanical linear actuators only pull (e.g. hoist, chain drive and belt drive) and others only push (e.g. cam actuator).

Mechanical actuators typically convert rotary motion of a control knob or handle into linear displacement via screws and/or gears to which the knob or handle is attached. A jackscrew or car jack is a familiar mechanical actuator. Another family of actuators are based on the segmented spindle. Rotation of the jack handle is converted mechanically into the linear motion of the jack head. Mechanical actuators are also frequently used in the field of lasers and optics to manipulate the position of linear stages, rotary stages, mirror mounts, goniometers and other positioning instruments. For accurate and repeatable positioning, index marks may be used on control knobs. Some actuators even include an encoder and digital position readout. These are similar to the adjustment knobs used on micrometers except that their purpose is position adjustment rather than position measurement.

### Hydraulic actuators

Hydraulic actuators or hydraulic cylinders typically involve a hollow cylinder having a piston inserted in it. The two sides of the piston are alternately pressurized/de-pressurized to achieve controlled precise linear displacement of the piston and in turn the entity connected to the piston. The physical linear displacement is only along the axis of the piston/cylinder. This design is based on the principles of hydraulics. A familiar example of a manually operated hydraulic actuator is a hydraulic car jack. Typically though, the term "hydraulic actuator" refers to a device controlled by a hydraulic pump.

### Pneumatic actuators

Pneumatic actuators, or pneumatic cylinders, are similar to hydraulic actuators except they use compressed gas to provide pressure instead of a liquid.

### Piezoelectric actuators

The piezoelectric effect is a property of certain materials in which application of a voltage to the material causes it to expand. Very high voltages correspond to only tiny expansions. As a result, piezoelectric actuators can achieve extremely fine positioning resolution, but also have a very short range of motion. In addition, piezoelectric materials exhibit hysteresis which makes it difficult to control their expansion in a repeatable manner.

**Electro-mechanical actuators**

A miniature electro-mechanical linear actuator where the lead nut is part of the motor. The lead screw does not rotate, so as the lead nut is rotated by the motor, the lead screw is extended or retracted.

Typical compact cylindrical linear electric actuator


Typical linear or rotary + linear electric actuator

Moving coil linear, rotary and linear + rotary actuators at work in various applications

Electro-mechanical actuators are similar to mechanical actuators except that the control knob or handle is replaced with an electric motor. Rotary motion of the motor is converted to linear displacement of the actuator. There are many designs of modern linear actuators and every company that manufactures them tends to have their own proprietary method. The following is a generalized description of a very simple electro-mechanical linear actuator.

## Simplified design

Typically, a rotary driver (e.g. electric motor) is mechanically connected to a lead screw so that the rotation of the electric motor will make the lead screw rotate. A lead screw has a continuous helical thread machined on its circumference running along the length (similar to the thread on a bolt). Threaded onto the lead screw is a lead nut or ball nut with corresponding helical threads. The nut is prevented from rotating with the lead screw (typically the nut interlocks with a non-rotating part of the actuator body). Therefore, when the lead screw is rotated, the nut will be driven along the threads. The direction of motion of the nut will depend on the direction of rotation of the lead screw. By connecting linkages to the nut, the motion can be converted to usable linear displacement. Most current actuators are built either for high speed, high force, or a compromise between the two. When considering an actuator for a particular application, the most important specifications are typically travel, speed, force, accuracy, and lifetime.

There are many types of motors that can be used in a linear actuator system. These include dc brush, dc brushless, stepper, or in some cases, even induction motors. It all depends on the application requirements and the loads the actuator is designed to move. For example, a linear actuator using an integral horsepower AC induction motor driving a lead screw can be used to actuate a large valve in a refinery. In this case, accuracy and move resolution down to a thousandth isn't needed, but high force and speed is. For electromechanical linear actuators used in laboratory instrumentation robotics, optical and laser equipment, or X-Y tables, fine resolution into the micron region and high accuracy may require the use of a fractional horsepower stepper motor linear actuator with a fine pitch lead screw. There are many variations in the electromechanical linear actuator system. It's critical to understand the design requirements and application constraints to know which one would be best.

## Principles

In the majority of linear actuator designs, the basic principle of operation is that of an inclined plane. The threads of a lead screw act as a continuous ramp that allows a small rotational force to be used over a long distance to accomplish movement of a large load over a short distance.

## Variations

Many variations on the basic design have been created. Most focus on providing general improvements such as a higher mechanical efficiency, speed, or load capacity. There is also a large engineering movement towards actuator miniaturization.

Most electro-mechanical designs incorporate a lead screw and lead nut. Some use a ball screw and ball nut. In either case the screw may be connected to a motor or manual control knob either directly or through a series of gears. Gears are typically used to allow a smaller (and weaker) motor spinning at a higher rpm to be geared down to provide the torque necessary to spin the screw under a heavier load than the motor would otherwise be capable of driving directly. Effectively this sacrifices actuator speed in favor of increased actuator thrust. In some applications the use of worm gear is common as this allow a smaller built in dimension still allowing great travel length.

A traveling-nut linear actuator has a motor that stays attached to one end of the lead screw (perhaps indirectly through a gear box), the motor spins the lead screw, and the lead nut is restrained from spinning so it travels up and down the lead screw.

A traveling-screw linear actuator has a lead screw that passes entirely through the motor. In a traveling-screw linear actuator, the motor "crawls" up and down a lead screw that is restrained from spinning—the only spinning parts are inside the motor, and may not even be visible from the outside.

Some lead screws have multiple "starts". This means that they have multiple threads alternating on the same shaft. One way of visualizing this is in comparison to the multiple color stripes on a candy cane. This allows for more adjustment between thread pitch and nut/screw thread contact area, which determines the extension speed and load carrying capacity (of the threads), respectively.

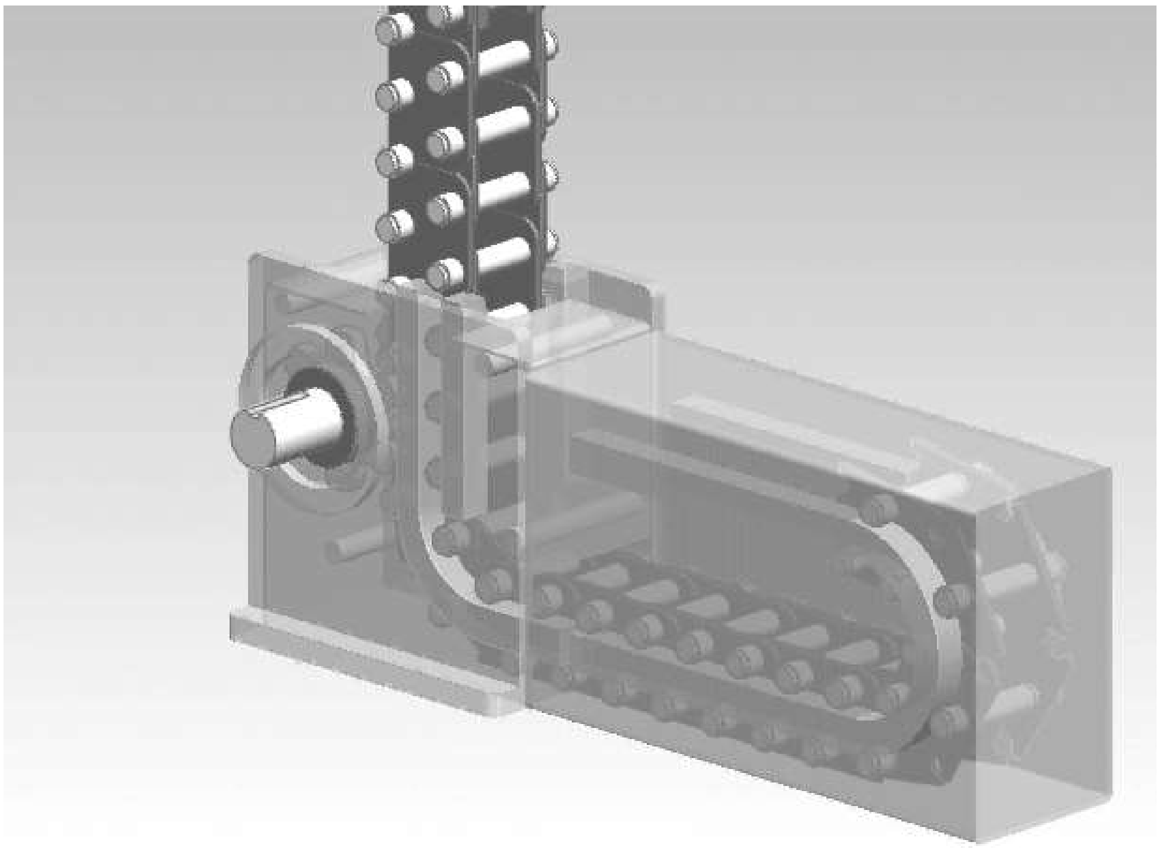
## Linear motors

A linear motor is essentially a rotary electric motor laid down on flat surface. Since the motor moves in a linear fashion to begin with, no lead screw is needed to convert rotary motion to linear. While high capacity is possible, the material and/or motor limitations on most designs are surpassed relatively quickly. Most linear motors have a low load capacity compared to other types of linear actuators.

### Wax motors

A wax motor typically uses an electric current to heat a block of wax causing it to expand. A plunger that bears on the wax is thus forced to move in a linear fashion.

### Telescoping linear actuator



Rigid chain actuator

Telescoping linear actuators are specialized linear actuators used where space restrictions or other requirements require. Their range of motion is many times greater than the unextended length of the actuating member.

A common form is made of concentric tubes of approximately equal length that extend and retract like sleeves, one inside the other, such as the telescopic cylinder.

Other more specialized telescoping actuators use actuating members that act as rigid linear shafts when extended, but break that line by folding, separating into pieces and/or uncoiling when retracted. Examples of telescoping linear actuators include:

- Helical band actuator
- Rigid belt actuator
- Rigid chain actuator
- Segmented spindle

# Advantages and disadvantages

| Actuator Type | Advantages | Disadvantages |
|---|---|---|
| Mechanical | Cheap. Repeatable. No power source required. Self contained. Identical behaviour extending or retracting. | Manual operation only. No automation. |
| Electro-mechanical | Cheap. Repeatable. Operation can be automated. Self-contained. Identical behaviour extending or retracting. DC or stepping motors. Position feedback possible. | Many moving parts prone to wear. |
| Linear motor | Simple design. Minimum of moving parts. High speeds possible. Self-contained. Identical behaviour extending or retracting. | Low force. |
| Piezoelectric | Very small motions possible. | Requires position feedback to be repeatable. Short travel. Low speed. High voltages required. Expensive. Good in compression only, not in tension. |
| Hydraulic | Very high forces possible. | Can leak. Requires position feedback for repeatability. External hydraulic pump required. Some designs good in compression only. |
| Pneumatic | Strong, light, simple, fast. | Precise position control impossible except at full stops |
| Wax motor | Smooth operation. | Not as reliable as other methods. |
| Segmented spindle | Very compact. Range of motion greater than length of actuator. | Both linear and rotary motion. |
| Moving coil | Force, position and speed are controllable and repeatable. Capable of high speeds and precise positioning. Linear, rotary, and linear + rotary actions possible. | Requires position feedback to be repeatable. |
| MICA (moving iron controllable actuator) | High force and controllable. Higher force and less losses than moving coils. Losses easy to dissipate. Electronic driver easy to design and set up. | Stroke limited to several millimeters, less linearity than moving coils |

## Chapter- 12

# Nanotube Nanomotor

A device generating linear or rotational motion using carbon nanotube(s) as the primary component, is termed a nanotube nanomotor. Nature already has some of the most efficient and powerful kinds of nanomotors. Some of these natural biological nanomotors have been re-engineered to serve desired purposes. However, such biological nanomotors are designed to work in specific environmental conditions (pH, liquid medium, sources of energy, etc). Man-made nanotube nanomotors on the other hand are significantly more robust and can operate in diverse environments including varied frequency, temperature, mediums and chemical environments. The vast differences in the dominant forces and criteria between macroscale and micro/nanoscale offer new avenues to construct tailor-made nanomotors. The various beneficial properties of carbon nanotubes makes them the most attractive material to base such nanomotors on.

## History

Just fifteen years after making the world's first micron sized motor Dr. Alex Zettl led his group at University of California at Berkeley to construct the very first nanotube nanomotor in 2003. A few concepts and models have been spun off ever since including the nanoactuator driven by a thermal gradient as well as the conceptual Electron Windmill, both of which were revealed in 2008.

## Size effects

### Electrostatic Forces

Coulomb's law states that the electrostatic force is inversely proportional to the square of the distance between two objects. Hence, as the distance is reduced to less than a few microns, a large force can be generated from seemingly small charges on two bodies. However electrostatic charge scales quadratically thereby resulting in the quadratic scaling of the electrostatic force as the following equations show:

$$Capacitance(C) = \frac{\varepsilon A}{d} \propto L$$

$$ElectrostaticField(E) \propto L^0$$

$$Voltage(V) = electrostaticfield * length = E * L \propto L$$

$$Charge = CV \propto L^2$$

$$ElectrostaticForce(F) = area * E^2 \propto L^2$$

Alternatively

$$ElectrostaticForce(F) = \frac{Q_1 * Q_2}{d^2} \propto \frac{L^2 * L^2}{L^2} \propto L^2$$

Despite the scaling nature of the electrostatic force it is one of the major mechanisms of sensing and actuation in the field of Micro-Electro-Mechanical Systems (MEMS) and is the backbone for the working mechanism of the first NEMS nanomotor. The quadratic scaling is alleviated by increasing the number of units generating the electrostatic force as seen in comb drives in many MEMS devices.

### Friction

Just as the electrostatic force, the frictional force scales quadratically with size:

$$FrictionalForce \propto L^2$$

Friction is an ever plaguing problem regardless of the scale of a device. It becomes all the more prominent when a device is scaled down. In the nano scale it can wreak havoc if not accounted for because the parts of a Nano-Electro-Mechanical-Systems (NEMS) device are sometimes only a few atoms thick. Furthermore such NEMS devices typically have a very large surface area-to-volume ratio. Surfaces in the nanoscale resemble a mountain range, where each peak corresponds to an atom or a molecule. Friction at the nanoscale is proportional to the number of atoms that interact between two surfaces. Hence, friction between perfectly smooth surfaces in the macroscale is actually similar to large rough objects rubbing against each other.

In the case of nanotube nanomotors however, the intershell friction in the multi walled nanotubes (MWNT) is remarkably small. Molecular dynamics studies show that, with the exception of small peaks, the frictional force remains almost negligible for all sliding velocities until a special sliding velocity is reached. Simulations relating the sliding velocity, induced rotation, inter-shell frictional force to the applied force provide explanations for the low inter-wall friction. Contrary to macroscale expectations the speed at which an inner tube travels within an outer tube does not follow a linear relationship with the applied force. Instead, the speed remains constant (as in a plateau) despite increasing applied force occasionally jumping in value to the next plateau. No real rotation is noticed in nonchiral inner tubes. In the case of chiral tubes a true rotation is noticed and the angular velocity also jumps to plateaus along with the jumps in the linear velocity. These plateaus and jumps can be explained as a natural outcome of frictional peaks for growing velocity, the stable (rising) side of the peak leading to a plateau, the dropping (unstable) side leading to a jump. These peaks occur due to parametric excitation of vibrational modes in the walls of the tubes due to the sliding of the inner tube. With the exception of small peaks, that correspond to the speed plateaus, the frictional force remains almost negligible for all sliding velocities until a special sliding velocity. These velocity plateaus correspond to the peaks in the frictional force. The sudden rise in sliding velocity is due to a resonance condition between a frequency that is dependent on the inter-tube corrugation period and particular phonon frequencies of the outer tube which happen to possess a group velocity approximately equal to the sliding velocity.

# First NEMS nanomotor

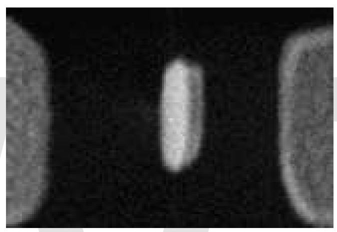


Figure 1.1: Image showing the metal rotor rotating around the MWNT axis

The first nanomotor can be thought of as a scaled down version of a comparable microelectromechanical systems (MEMS) motor. As Figure 1.1 and Figure 1.2 show, the nanoactuator consists of a gold plate rotor, rotating about the axis of a multi-walled nanotube (MWNT). The ends of the MWNT rest on a $SiO_2$ layer which form the two electrodes at the contact points. Three fixed stator electrodes (two visible 'in-plane' stators and one 'gate' stator buried beneath the surface) surround the rotor assembly. Four independent voltage signals (one to the rotor and one to each stators) are applied to control the position, velocity and direction of rotation. Empirical angular velocities recorded provide a lower bound of 17 Hz (although capable of operating at much higher frequencies) during complete rotations.

## Fabrication

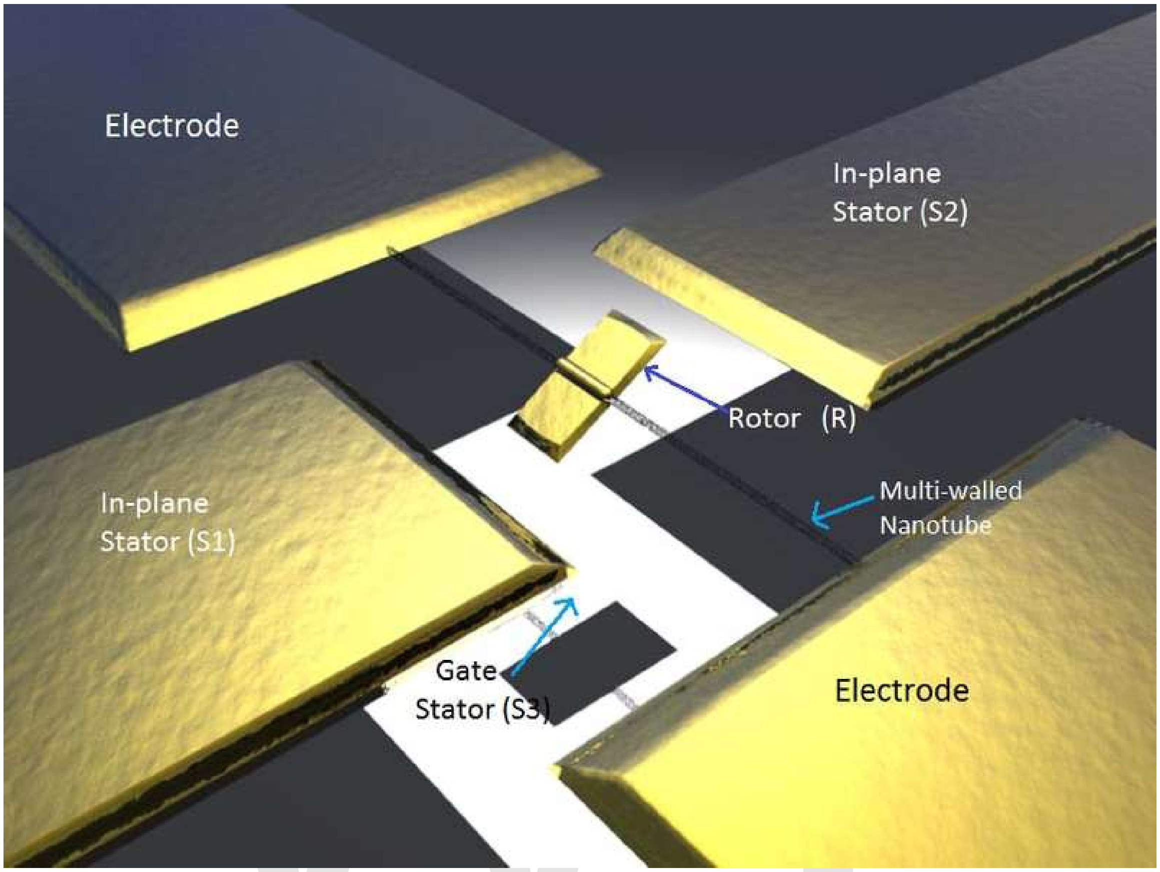



Figure 1.2: Schematic showing basic layout of the nanomotor

The MWNTs are synthesized by the arc-discharge technique, suspended in 1,2-dichlorobenzene and deposited on degenerately doped silicon substrates with a 1µ*m* of $SiO_2$. The MWNT can be aligned according to pre-made markings on the substrate by using an atomic force microscope (AFM) or a Scanning electron microscope. The rotor, electrodes and the 'in-plane' stators are patterned using electron beam lithography using an appropriately masked photo-resist. Gold with a chromium adhesion layer is thermally evaporated, lifted off in acetone and then annealed at 400°C to ensure better electrical and mechanical contact with the MWNT. The rotor measures 250-500 nm on a side. An HF etch is then used to remove sufficient thickness (500 nm of $SiO_2$) of the substrate to make room for the rotor when it rotates. The Si substrate serves as the gate stator. The MWNT at this point displays a very high torsional spring constant ($10^{-15}$ to $10^{-13}$ N m with resonant frequencies in the tens of megahertz), hence, preventing large angular displacements. To overcome this, one or more outer MWNT shells are compromised or removed in the region between the anchors and the rotor plate. One simple way to accomplish this is by successively applying very large stator voltages (around 80 V d.c) that cause mechanical fatigue and eventually shear the outer shells of the MWNT. An alternative method involves the reduction of the outermost MWNT tubes to smaller, wider concentric nanotubes beneath the rotor plate.

The smaller nanotube(s) are fabricated using the Electrical driven vaporization (EDV) which is a variant of the electrical-breakdown technique. Passing current between the two electrodes typically results in failure of the outermost shell only on one side of the nanotube. Current is therefore passed between one electrode and the center of the MWNT which results in the failure of the outermost shell between this electrode and the center. The process is repeated on the opposite side to result in the formation of the short concentric nanotube that behaves like a low friction bearing along the longer tube.

## Arrays of Nanoactuators

Due to the miniscular magnitude of output generated by a single nanoactuator the necessity to use arrays of such actuators to accomplish a higher task comes into picture. Conventional methods like chemical vapor deposition (CVD) allow the exact placement of nanotubes by growing them directly on the substrate. However, such methods are unable to produce very high qualities of MWNT. Moreover, CVD is a high temperature process that would severely limit the compatibility with other materials in the system. A Si substrate is coated with electron beam resist and soaked in acetone to leave only a thin polymer layer. The substrate is selectively exposed to an low energy electron beam of a scanning electron microscope (SEM) that activates the adhesive properties of the polymer later. This forms the basis for the targeting method. The alignment method exploits the surface velocity obtained by a fluid as it flows off a spinning substrate. MWNTs are suspended in orthodicholrobenzene (ODCB) by ultrasonication in a aquasonic bath that separates most MWNT bundles into individual MWNTs. Drops of this suspension are then pipetted one by one onto the center of a silicon substrate mounted on a spin coater rotating at 3000 rpm. Each subsequent drop of the suspension is pipetted only after the previous drop has completely dried to ensure larger density and better alignment of the MWNTs (90% of the MWNTs over 1 μm long lie within 1°). Standard electron beam lithography is used to pattern the remaining components of the nanoactuators.

## Arc-Discharge Evaporation technique

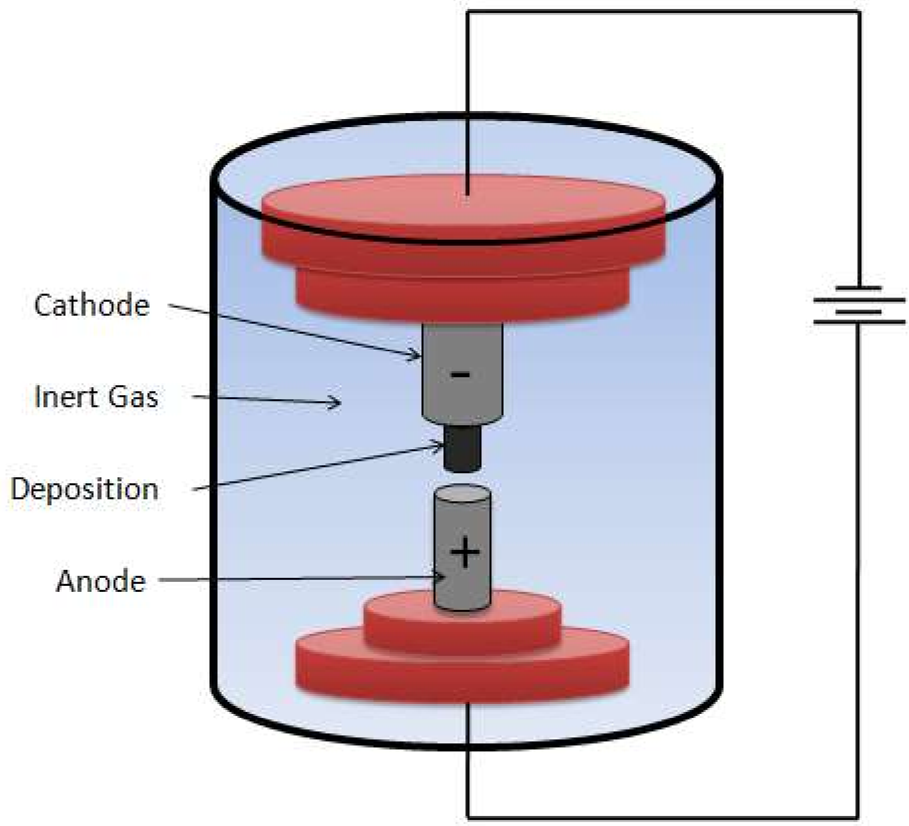



Figure 1.3: Cartoon showing the basic experimental setup for the Arc-Discharge technique of large scale carbon nanotube synthesis

This technique is a variant of the standard arc-discharge technique used for the synthesis of fullerenes in an innert gas atmosphere. As Figure 1.3 shows, the experiment is carried out in a reaction vessel containing an inert gas such as helium, argon, etc flowing at a constant pressure. A potential of around 18V is applied across two graphite electrodes (diameters of the anode and cathode are 6 mm and 9 mm) separated by a short distance of usually 1–4 mm within this chamber. The amount of current (usually 50-100 A) passed through the electrodes to ensure nanotube formation depends on the dimensions of the electrodes, separation distance and the inert gas used. As a result, carbon atoms are ejected from the anode and are deposited onto the cathode hence shrinking the mass of the anode and increasing the mass of the cathode. The black carbonaceous deposit (a mixture of nanoparticles and nanotubes in a ratio of 1:2) is seen growing on the inside of the cathode while a hard grey metallic shell forms on the outside. The total yield of nanotubes as a proportion of starting graphitic material peaks at a pressure of 500 torr at which point 75% of graphite rod consumed is converted to nanotubes. The nanotubes formed range from 2 to 20 nm in diameter and few to several microns in length. There are several advantages of choosing this method over the other techniques such as laser ablation and chemical vapor deposition such as fewer structural defects (due to high growth temperature), better electrical, mechanical and thermal properties, high production rates (several hundred mg in ten minutes), etc.

### Electrical-breakdown technique

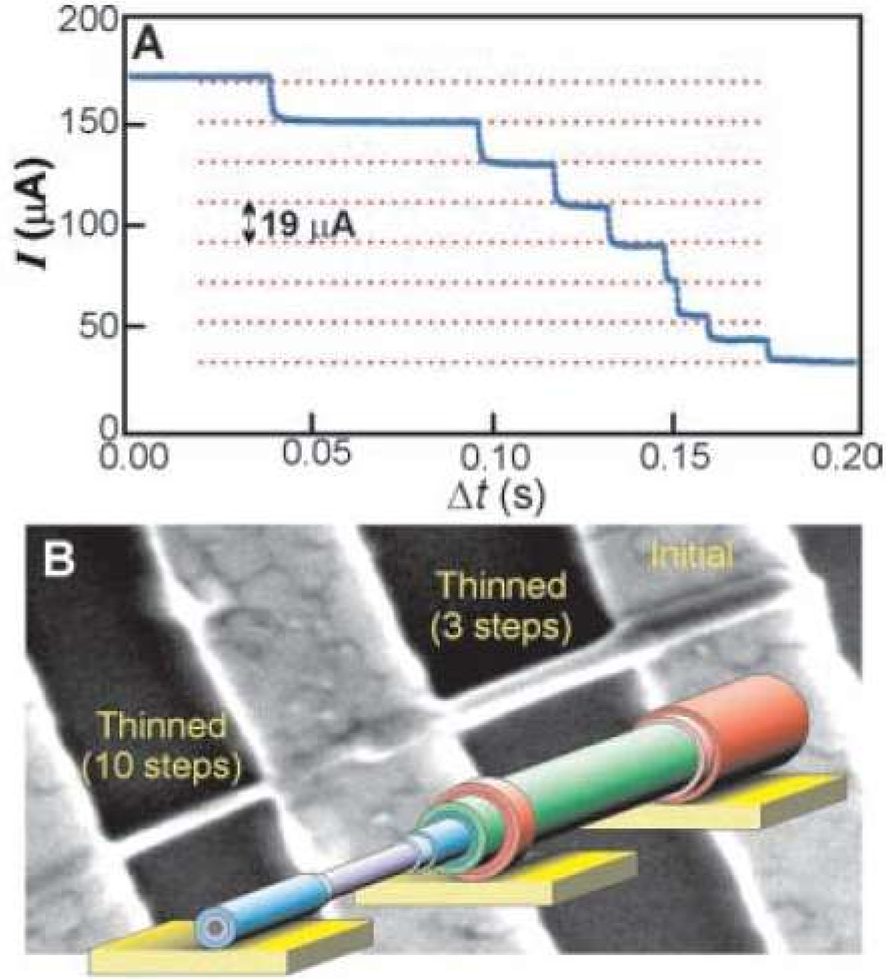



Figure 1.4: (A) Graph showing remarkably discrete, constant drops in conductance for the removal of each subsequent carbon shell under constant voltage (B) Images of partially broken MWNTs show clear thinning, with a decrease in radius equal to the intershell spacing (0.34 nm) times the number of completed breakdown steps. The two segments of this sample were independently thinned by 3 and 10 shells, as depicted by the color overlays

Large-scale synthesis of carbon nanotubes typically results in a randomly varied proportion of different types of carbon nanotubes. Some may be semiconducting while others may be metallic in their electrical properties. Most applications require the use of such specific types of nanotubes. Electrical-breakdown technique provides a means for separating and selecting desired type of nanotubes. Carbon nanotubes are known to withstand very large current densities up to $10^9$ A/cm$^2$ partly due to the strong sigma bonds between carbon atoms. However, at sufficiently high currents the nanotubes fail primarily due to rapid oxidation of the outermost shell. This results in a partial conductance drop that becomes apparent within a few seconds. Applying an increased bias displays multiple independent and stepwise drops in conductance (figure 1.4) resulting from the sequential failure of carbon shells. Current in a MWNT typically travels in the outermost shell due the direct contact between this shell and the electrodes. This controlled destruction of shells without affecting disturbing inner layers of MWNTs permits the effective separation of the nanotubes.

## Principle

The rotor is made to rotate using electrostatic actuation. An out-of-phase common frequency sinusoidal voltages to two in-plane stators $S_1$, $S_2$, a doubled frequency voltage signal to the gate stator $S_3$ and a dc offset voltage to the rotor plate R are applied as shown below:

$$S_1 = V_0 sin(\omega t)$$

$$S_2 = V_0 sin(\omega t - \pi)$$

$$S_3 = V_0 sin(2\omega t + \frac{\pi}{2})$$

$$R = - V_0$$

By the sequential application of these asymmetrical stator voltages (less than 5 V) the rotor plate can be drawn to successive stators hence making the plate complete rotations. The high proximity between the stators and the rotor plate is one reason why a large force is not required for electrostatic actuation. Reversing the bias causes the rotor to rotate in the opposite direction as expected.

## Applications

- The rotating metal plate could serve as a mirror for ultra-high-density optical sweeping and switching devices as the plate is at the limit of visible light focusing. An array of such actuators, each serving as a high frequency mechanical filter, could be used for parallel signal processing in telecommunications.
- The plate could serve as a paddle for inducing or detecting fluid motion in microfluidic applications. It could serve as a bio-mechanical element in biological systems, a gated catalyst in wet chemistry reactions or as a general sensor element.
- A charged oscillating metal plate could be used as a transmitter of electromagnetic radiation.

# Thermal Gradient driven Nanotube actuators

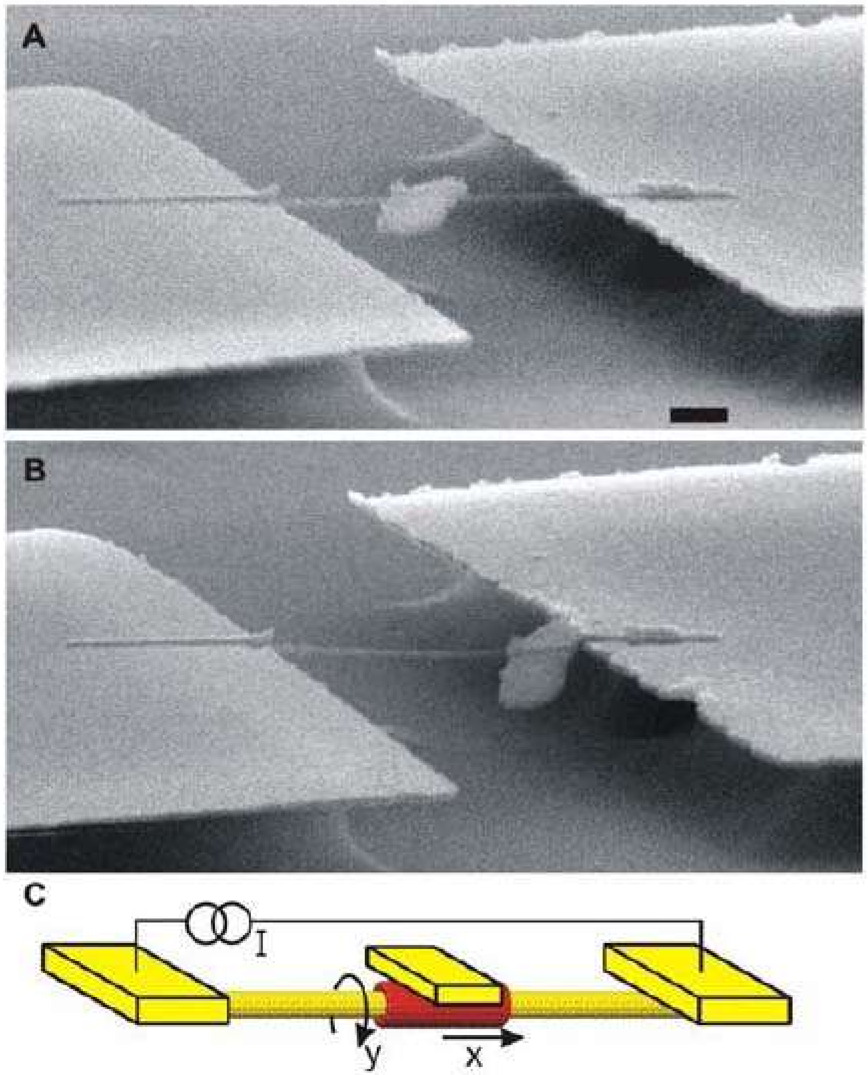



Figure 2.1: Thermal gradient driven nanomotor. (A & B): Scanning Electron microscope (SEM) images of experimental setup. (C) Schematic of the nanomotor also displaying degrees of freedom

The nanoactuator, as shown in Figure 2.1 comprises two electrodes connected via a long MWNT. A gold plate acts as the cargo and is attached to a shorter and wider concentric nanotube. The cargo moves towards the cooler electrode (Figure 2.2) due to the thermal gradient in the longer nanotube induced by the high current that is passed through it. The maximum velocity was approximated to $1\mu m / sec$ which is comparable to the speeds attained by kinesin biomotors.

## Fabrication

The MWNT are fabricated using the standard arc-discharge evaporation process and deposited on an oxidized silicon substrate. The gold plate in the center of the MWNT is patterned using electron-beam lithography and Cr/Au evaporation. During the same process, the electrodes are attached to the nanotube. Finally, electrical-breakdown technique is used to selectively remove a few outer walls of the MWNT. Just as the nanoactuator from the Zettl group, this enables low friction rotation and translation of the shorter nanotube along the axis of the longer tube. The application of the electrical-breakdown technique does not result in the removal of the tube(s) below the cargo. This might be because the metal cargo absorbs the heat generated in the portion of the tube in its immediate vicinity hence delaying or possibly even preventing tube oxidation in this part.

**Principle**

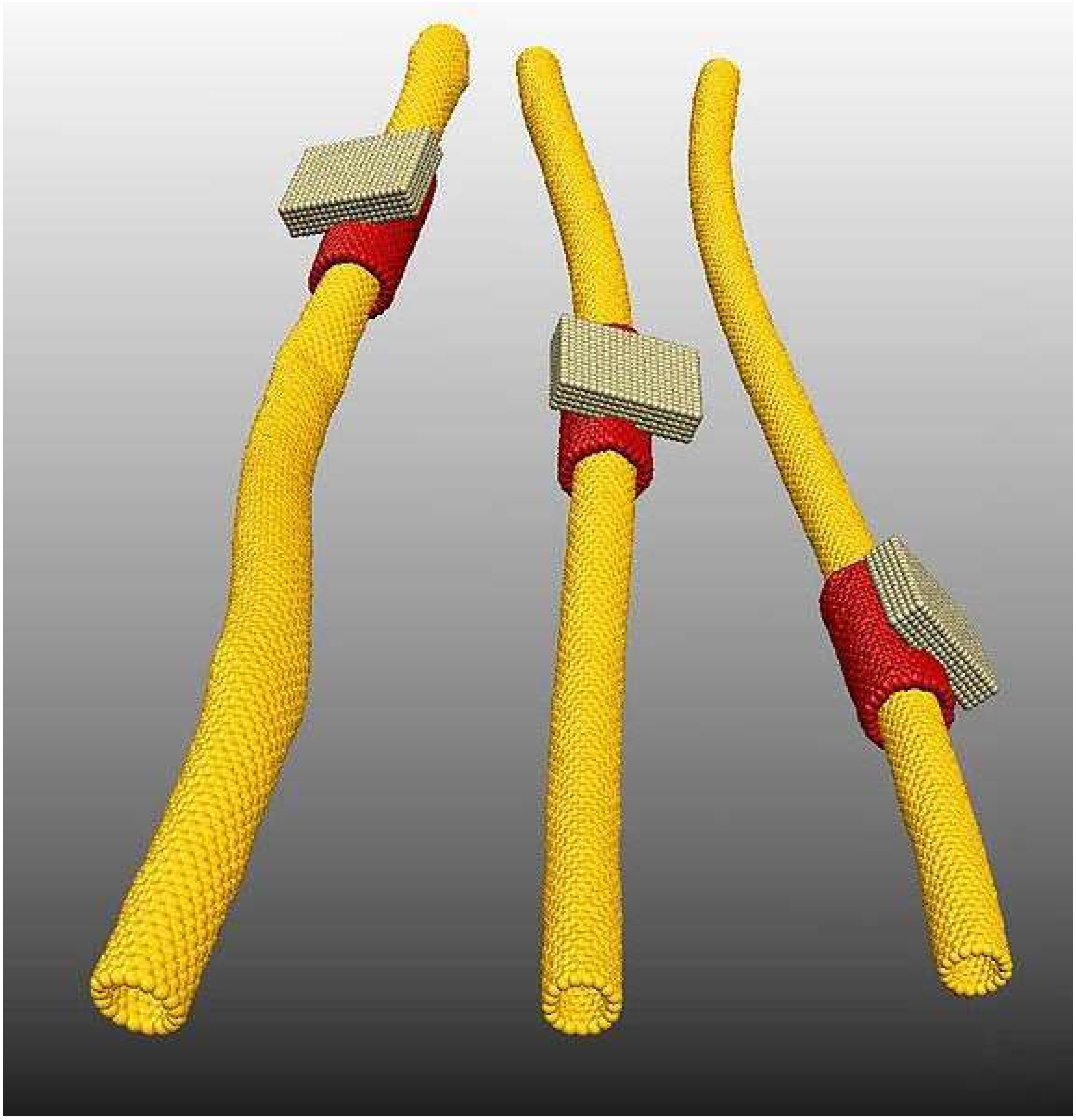


Figure 2.2: Motion of shorter nanotubes (red) along the longer tubes (yellow) from the hotter(top) section of the nanotube to the cooler (bottom) section of the nanotube carrying the metal cargo (gray)

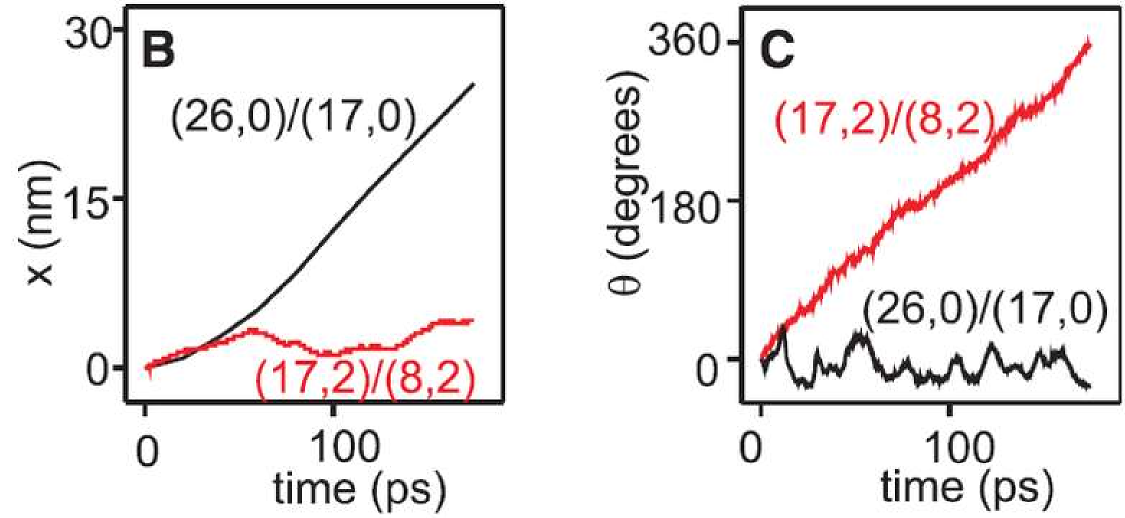



Figure 2.3: Degree of translational and rotation are dependent on the chiralities of the two nanotubes

The interaction between the longer and shorter tubes generates an energy surface that confines the motion to specific tracks - translation and rotation. The degree of translational and rotational motion of the shorter tube are highly dependent on the chiralities of the two tubes as shown in Figure 2.3. Motion in the nanoactuator displayed a proclivity of the shorter tube to follow a path of minimum energy. This path could either have a roughly constant energy or have a series of barriers. In the former case, friction and vibrational motion of atoms can be neglected whereas a stepwise motion is expected in the latter scenario.

## Stepwise motion

The stepwise motion can be explained by the existence of periodic energy barriers for relative motion between the longer and shorter tubes. For a given pair of nanotubes, the ratio of the step in rotation to the step in translation is typically a constant, the value of which depends on the chirality of the nanotubes. The energy of such barriers could be estimated from the temperature in the nanotube, a lower bound for which can be estimated as the melting temperature of gold (1300 K) by noting that the gold plate melts (Figure 2.4) to form a spherical structure as current is passed through the nanomotor. The motion rate Γ can be written as a function of the attempt frequency ω, the Boltzmann's constant *k*, and temperature *T* as:

$$\Gamma = \frac{\omega}{2\pi} e^{\frac{-\Delta E}{kT}}$$

Taking $\Gamma \approx 1Hz$, using the approximation:

$$\omega = \sqrt{\frac{\Delta E}{m a_0^2}}$$

where *m* is the mass of the cargo and $a_0^2$ represents the contact area, the barrier height is obtained to be $\approx 17 \mu eV$ per atom.

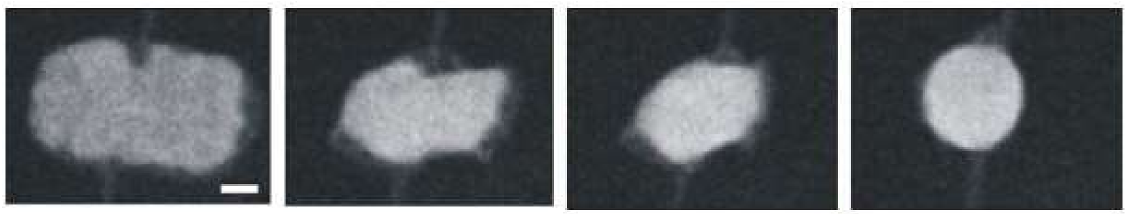


Figure 2.4: SEM images show the transformation of gold plate (left) into a ball (right) due to very high temperatures

## Mechanism for actuation

Many proposals were made to explain the driving mechanism behind the nanoactuator. The high current (0.1 mA) required to drive the actuator is likely to cause sufficient dissipation to clean the surface of contaminants; hence, ruling out the possibility of contaminants playing a major role. The possibility of electromigration, where the electrons move atomic impurities via momentum transfer due to collisions, was also ruled out because the reversal of the current direction did not affect the direction of displacement. Similarly, rotational motion could not have been caused by an induced magnetic field due to the current passing through the nanotube because the rotation could either be left or right-handed depending on the device. Stray electric field effect could not be the driving factor because the metal plate staid immobile for high resistive devices even under a large applied potential. The thermal gradient in the nanotube provides the best explanation for the driving mechanism.

## Thermal gradient induced motion

The induced motion of the shorter nanotube is explained as the reverse of the heat dissipation that occurs in friction wherein the sliding of two objects in contact results in the dissipation of some of the kinetic energy as phononic excitations caused by the interface corrugation. The presence of a thermal gradient in a nanotube causes a net current of phononic excitations traveling from the hotter region to the cooler region. The interaction of these phononic excitations with mobile elements (the carbon atoms in the shorter nanotube) causes the motion of the shorter nanotube. This explains why the shorter nanotube moves towards the cooler electrode. Changing the direction of the current has no effect on the shape of thermal gradient in the longer nanotube. Hence, direction of the movement of the cargo is independent of the direction of the bias applied. The direct dependence of the velocity of the cargo to the temperature of the nanotube is inferred from the fact that the velocity of the cargo decreases exponentially as the distance from the midpoint of the long nanotube increases.

### Shortcomings

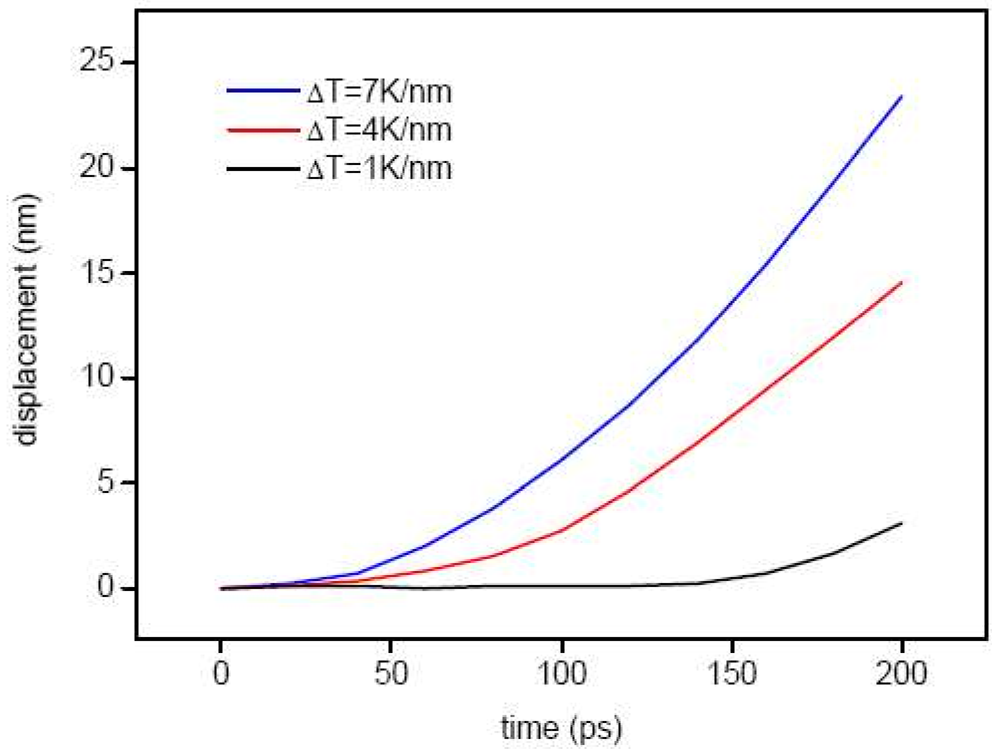



Figure 2.5: Graph demonstrating the direct relationship between the thermal gradient and the displacement of the shorter tube / cargo

The temperatures and the thermal gradient that the MWNT are subjected to are very high. On one hand, the high thermal gradient seems to have a highly detrimental effect on the lifetime of such nanoactuators. On the other hand, experiments show that the displacement of the shorter tube is directly proportional to the thermal gradient (see Figure 2.5). Therefore, a compromise needs to be reached to optimize the thermal gradient. The dimensions of movable nanotube is directly related to the energy barrier height. Although the current model excites multiple phonon modes, selective phonon mode excitation would enable lowering the phonon bath temperature.

### Applications

1. Pharmaceutical / Nanofluidic - Thermal gradient could be used to drive fluids within the nanotubes or in nanofluidic devices as well as for drug delivery by nanosyringes.

2. Running bio engineered nanopores using heat generated from adenosine triphosphate (ATP) molecules.

# Electron Windmill

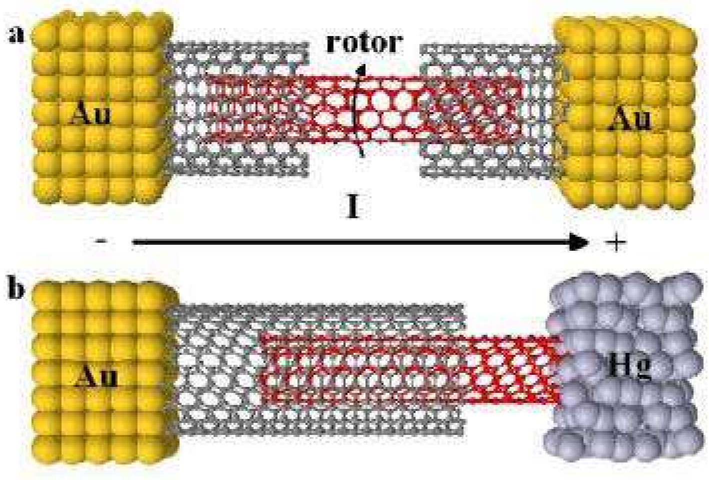



Figure 3.1: MWNT nanomotor(A) and nanodrill(B).

## Structure

As figure 3.1 shows, the nanomotor consists of a double-walled CNT (DWNT) formed from an achiral (18,0) outer tube clamped to external gold electrodes and a narrower chiral (6,4) inner tube. The central portion of the outer tube is removed using the electrical-breakdown technique to expose the free-to-rotate, inner tube. The nanodrill also comprises an achiral outer nanotube attached to a gold electrode but the inner tube is connected to a mercury bath.

## Principle

Conventional nanotube nanomotors make use of static forces that include elastic, electrostatic, friction and van der Waals forces. The electron windmill model makes use of a new "electron-turbine" drive mechanism that obviates that need for metallic plates and gates that the above nanoactuators require. When a dc voltage is applied between the electrodes, a "wind" of electrons is produced from left to right. The incident electron flux in the outer achiral tube initially possesses zero angular momentum, but acquires a finite angular momentum after interacting with the inner chiral tube. By Newton's third law, this flux produces a tangential force (hence a torque) on the inner nanotube causing it to rotate hence giving this model the name - "electron windmill". For moderate voltages, the tangential force produced by the electron wind is much greatly exceed the associated frictional forces.

## Applications

Some of the main applications of the electron windmill include:

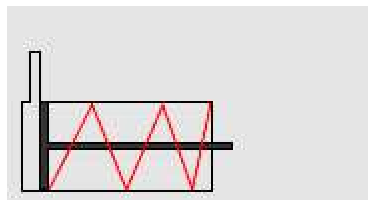
- A voltage pulse could cause the inner element to rotate at a calculated angle hence making the device behave as a switch or a nanoscale memory element.
- Modification of the electron windmill to construct a nanofluidic pump by replacing the electrical contacts with reservoirs of atoms or molecules under the influence of an applied pressure difference.

## Further developments

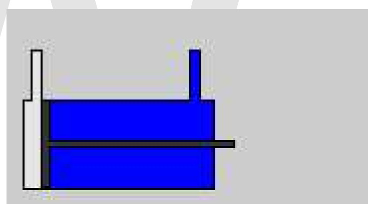
Although it has been demonstrated that synthetic nanotube nanoactuators can be built and are highly controllable, they are currently are not able to match the biological and macroscale nanomotors in terms of efficiency and scalability to accomplish higher tasks. However, further research, the electron windmill for example in this area promises to overcome this problem very soon.

# Chapter- 13

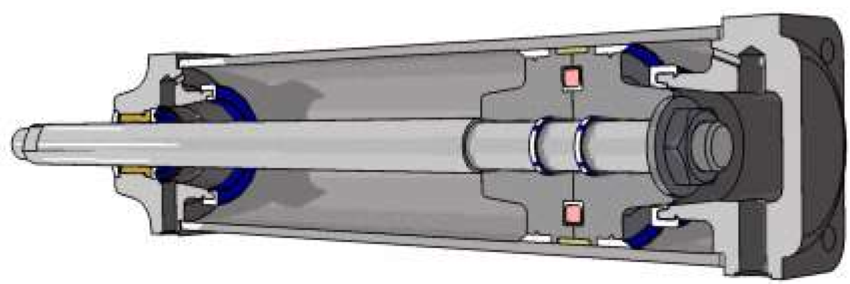
# Pneumatic Cylinder



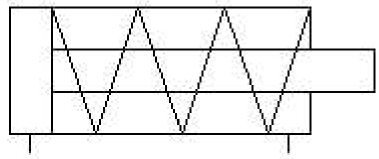
Operation diagram of a single acting cylinder. The spring (red) can also be outside the cylinder, attached to the item being moved.



Operation diagram of a double acting cylinder



3D image of pneumatic cylinder (CAD)



Schematic symbol for pneumatic cylinder with spring return

**Pneumatic cylinders** (sometimes known as **air cylinders**) are mechanical devices which produce force, often in combination with movement, and are powered by compressed gas (typically air).

To perform their function, pneumatic cylinders impart a force by converting the potential energy of compressed gas into kinetic energy. This is achieved by the compressed gas being able to expand, without external energy input, which itself occurs due to the pressure gradient established by the compressed gas being at a greater pressure than the atmospheric pressure. This air expansion forces a piston to move in the desired direction. The piston is a disc or cylinder, and the piston rod transfers the force it develops to the object to be moved.

### A note about popular terminology

At least in the USA, popular usage sometimes refers to the whole assembly of cylinder, piston, and piston rod (or more) collectively as a "piston", which is incorrect. See, for instance, "Hydraulic piston raises the table from 19 (in.) to 26 (in.)" Marine Tables, Inc. (Select item 3 of 8, near the bottom.)

# Operation

### General

Once actuated, compressed air enters into the tube at one end of the piston and, hence, imparts force on the piston. Consequently, the piston becomes displaced (moved) by the compressed air expanding in an attempt to reach atmospheric pressure.

### Fail safe mechanisms

Pneumatic systems are often found in settings where even rare and brief system failure is unacceptable. In such situations locks can sometimes serve as a safety mechanism in case of loss of air supply (or its pressure falling) and, thus, remedy or abate any damage arising in such a situation. Due to the leakage of air from input or output reduces the pressure and so the desired output.

# Types

Although pneumatic cylinders will vary in appearance, size and function, they generally fall into one of the specific categories shown below. However there are also numerous other types of pneumatic cylinder available, many of which are designed to fulfill specific and specialised functions.

## Single acting cylinders

Single acting cylinders (SAC) use the pressure imparted by compressed air to create a driving force in one direction (usually out), and a spring to return to the "home" position.

## Double acting cylinders

Double Acting Cylinders (DAC) use the force of air to move in both extend and retract strokes. They have two ports to allow air in, one for outstroke and one for instroke.

## Other types

Although SACs and DACs are the most common types of pneumatic cylinder, the following types are not particularly rare:

- Rotary air cylinders: actuators that use air to impart a rotary motion
- Rodless air cylinders: These have no piston rod. They are actuators that use a mechanical or magnetic coupling to impart force, typically to a table or other body that moves along the length of the cylinder body, but does not extend beyond it.

## Rodless cylinders

Some rodless types have a slot in the wall of the cylinder. That slot is closed off for much of its length by two flexible metal sealing bands. The inner one prevents air from escaping, while the outer one protects the slot and inner band. The piston is actually a pair of them, part of a comparatively long assembly. They seal to the bore and inner band at both ends of the assembly. Between the individual pistons, however, are camming surfaces that "peel off" the bands as the whole sliding assembly moves toward the sealed volume, and "replace" them as the assembly moves away from the other end. Between the camming surfaces is part of the moving assembly that protrudes through the slot to move the load. Of course, this means that the region where the sealing bands are not in contact is at atmospheric pressure.

Another type has cables (or a single cable) extending from both (or one) end[s] of the cylinder. The cables are jacketed in plastic (nylon, in those referred to), which provides a smooth surface that permits sealing the cables where they pass through the ends of the cylinder. Of course, a single cable has to be kept in tension.

Still others have magnets inside the cylinder, part of the piston assembly, that pull along magnets outside the cylinder wall. The latter are carried by the actuator that moves the load. The cylinder wall is thin, to ensure that the inner and outer magnets are near each other. Multiple modern high-flux magnet groups transmit force without disengaging or excessive resilience.

1. ^ , (Catalog, 7.4 MB) Diagrams that show the principle are on Pages 6 and 7 (facing pair; it's worth configuring your reader). Only one piston is shown in the cutaway; the other is hidden; it is symmetrical, but reversed. Parker/Origa also makes similar cylinders with sealing bands.
2. ^ , Cable-type rodless cylinders
3. ^ , Commercial magnetically-coupled rodless cylinders

### Sizes

Air cylinders are available in a variety of sizes and can typically range from a small 2.5 mm air cylinder, which might be used for picking up a small transistor or other electronic component, to 400 mm diameter air cylinders which would impart enough force to lift a car. Some pneumatic cylinders reach 1000 mm in diameter, and are used in place of hydraulic cylinders for special circumstances where leaking hydraulic oil could impose an extreme hazard.

## Pressure, radius, area and force relationships

Although the diameter of the piston and the force exerted by a cylinder are related, they are not directly proportional to one another. Additionally, the typical mathematical relationship between the two assumes that the air supply does not become saturated. Due to the effective cross sectional area reduced by the area of the piston rod, the instroke force is less than the outstroke force when both are powered pneumatically and by same supply of compressed gas.

The relationship, between force on outstroke, pressure and radius, is as follows:

$$F = p(\pi r^2)$$

Cord

This is derived from the relationship, between force, pressure and **effective** *cross-sectional area*, which is:

F = p A\,

With the same symbolic notation of variables as above, but also *A* represents the effective cross sectional area.

On instroke, the same relationship between force exerted, pressure and *effective cross sectional area* applies as discussed above for outstroke. However, since the cross sectional area is less than the piston area the relationship between force, pressure and *radius* is different. The calculation isn't more complicated though, since the effective cross sectional area is merely that of the piston less that of the piston rod.

For instroke, therefore, the relationship between force exerted, pressure, radius of the piston, and radius of the piston rod, is as follows:

$$F = p(\pi r_1^2 - \pi r_2^2) = p\pi(r_1^2 - r_2^2)$$

Where:

$F$ represents the force exerted
$r_1$ represents the radius of the piston
$r_2$ represents the radius of the piston rod
$\pi$ is pi, approximately equal to 3.14159.

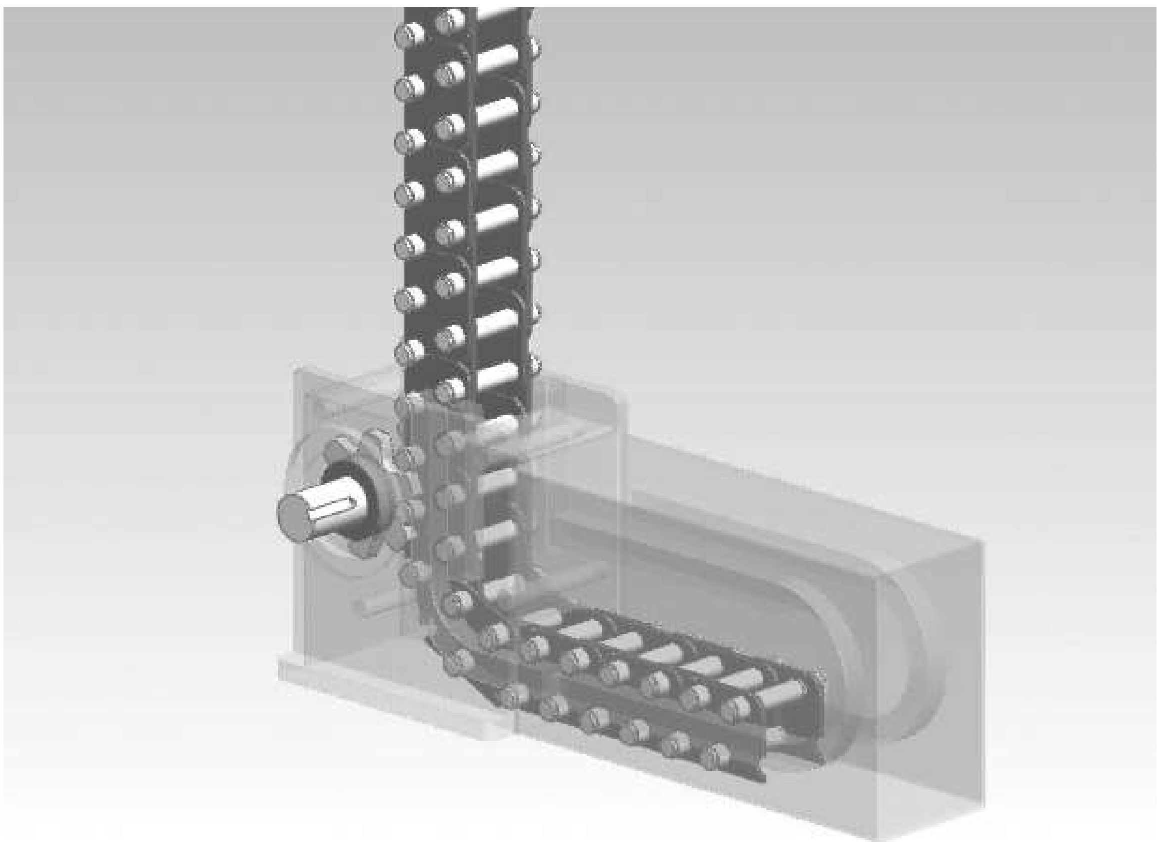
## Materials

The pneumatic cylinders designed for educational use typically have transparent outer sleeves (often plexiglass), so students can see the piston moving inside.

The pnuematic cylinders designed for cleanroom applications often use lubricant-free Pyrex glass pistons sliding inside graphite sleeves.

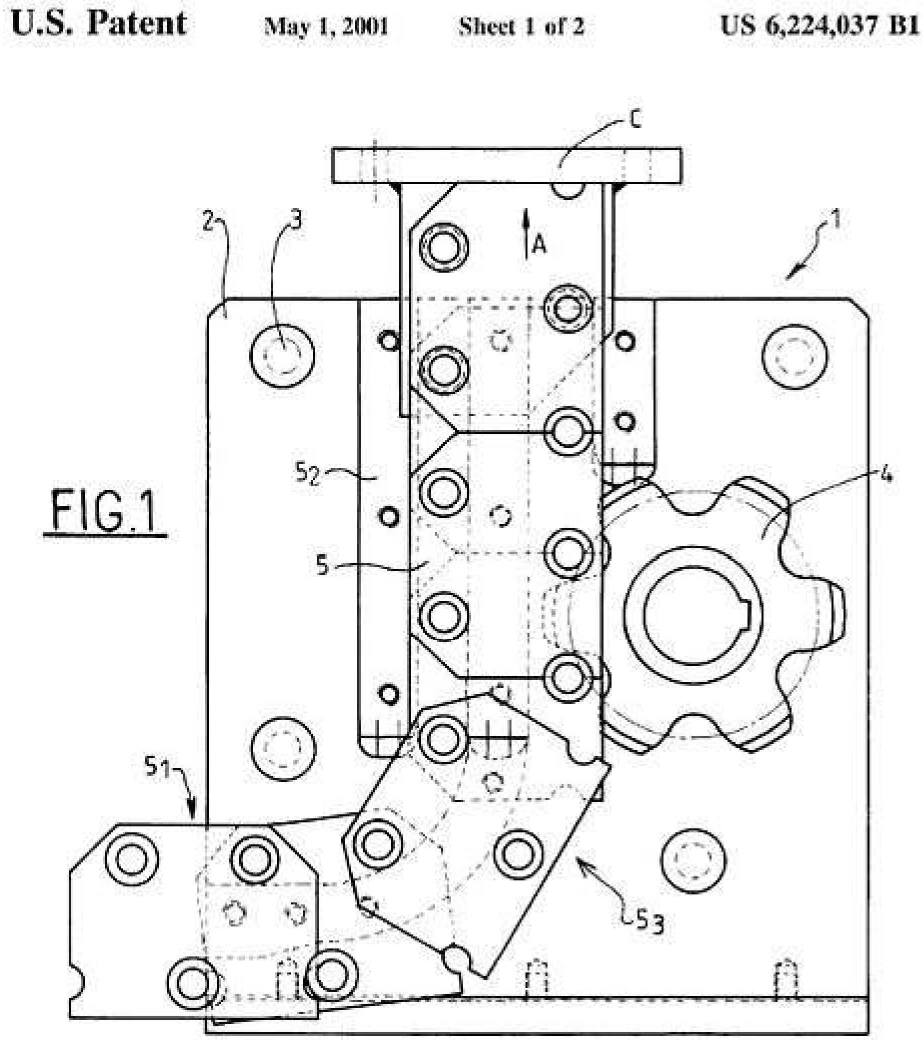
# Chapter- 14

# Rigid Chain Actuator



A **rigid chain actuator**, known variously as a **linear chain actuator**, **push-pull chain actuator**, **electric chain actuator** or **column-forming chain actuator**, is a specialized mechanical linear actuator used in window operating, push-pull material handling and lift applications. The actuator is a chain and pinion device that forms an articulated telescoping member to transmit traction and thrust. High-capacity rigid chain lifting columns (jacks) can move dynamic loads exceeding 10 tonnes (US 20,000 pounds) over more than 7 metres (20 feet) of travel.

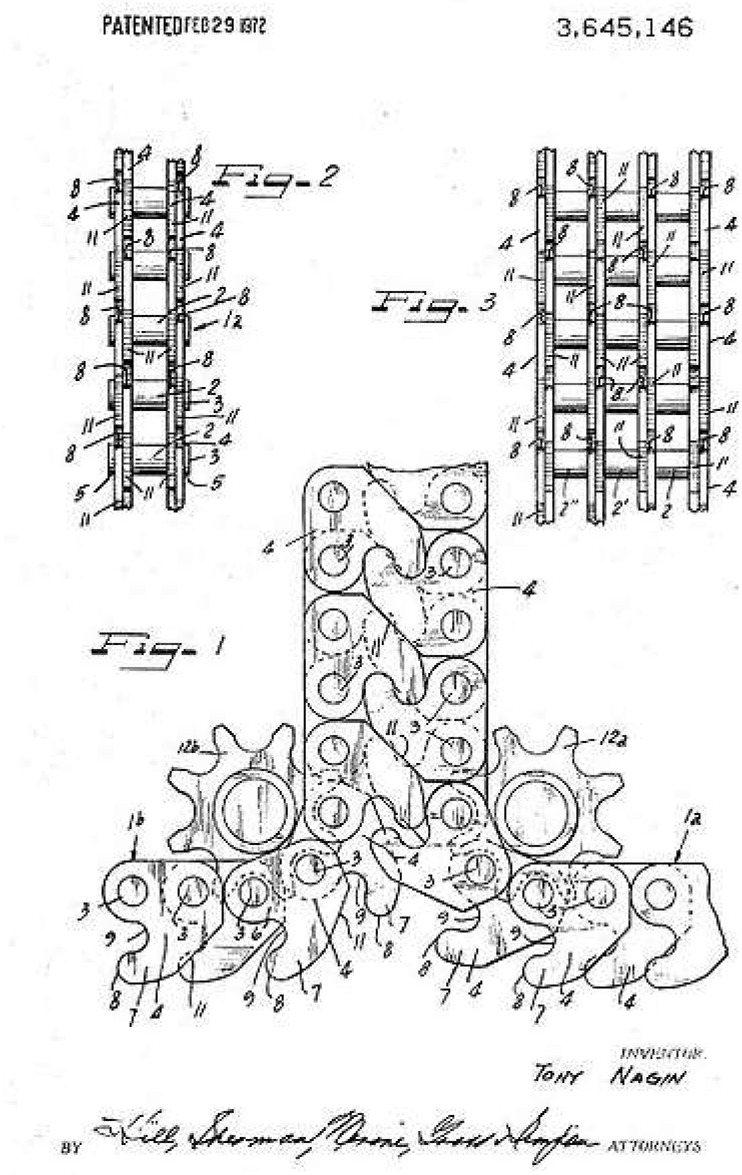
## Principle of operation




Patent Drawing for Serapid LinkLift Rigid Chain Actuator (2001).

Rigid chain actuators function as rack and pinion linear actuators that use articulated racks. Rigid chain actuators use limited-articulation chains, usually resembling a roller chain, that engage with pinions mounted on a drive shaft within a housing. The links of the actuating member, the “rigid chain”, are articulated in a manner that they deflect from a straight line to one side only. As the pinions spin, the links of the chain are rotated 90 degrees through the housing, which guides and locks the chain into a rigid linear form effective at resisting tension and compression (buckling). Because the actuating member can fold on itself, it can be stored relatively compactly in a storage magazine, either in an overlapping or coiled arrangement. Rigid chain actuators are generally driven by electric motors. Most rigid chains are manufactured from steel.

## Use




Patent Drawing for Interlocking Rigid Chain Actuator (1972).

Modified roller chain has been used extensively in material handling equipment, but could only be used in push-pull applications when a continuous loop of chain was used (with the exception of chain encapsulated in a guide channel). The development of efficient rigid chain actuators broadened the use of chain actuation for industrial applications. Small scale rigid chain actuators are used as building hardware, incorporated into windows, door and hatches as motorized open/close mechanisms. Rigid chain actuators are also used as the lifting columns in performing arts facilities, incorporated in stage, orchestra and seating platform lift systems.

Increasingly rigid chain systems are being incorporated into scissor lift tables or platform lifts as the method of actuation, replacing hydraulic cylinders. They are also used for production line automation and die changing.

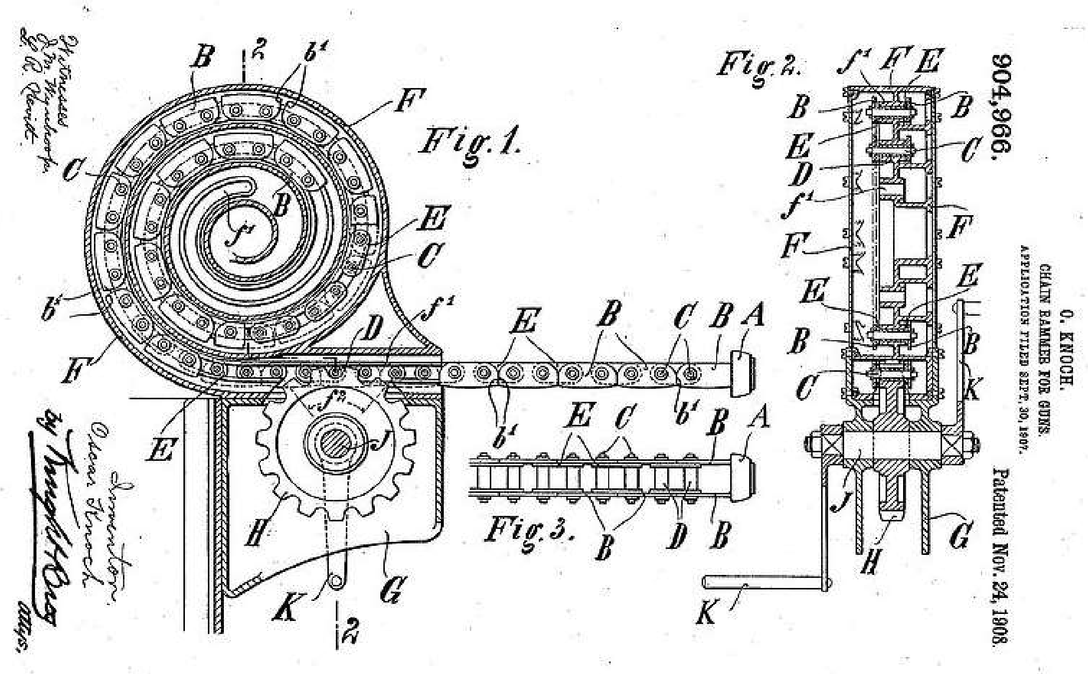
# Types

The primary distinction between types of rigid chain actuator is whether the actuating member is formed from a single chain or from a pair of interlocking chains in a back-to-back arrangement, like a zipper. Interlocking chain actuators have the advantages over single-chain actuators of improved resistance to buckling and that the actuating member does not require lateral restraint at its leading end in order to resist a modicum of transverse loads on any edge of the member. For example, it may function as a relatively stable telescoping pole.

The design of the chain varies significantly depending on application and manufacturer. Variants have been designed to, among other things:

- Simplify manufacture
- Reduce friction and maintenance
- Limit size and weight
- Increase speed, travel, capacity, efficiency and stability

## Development




Patent Drawing for Chain Rammer (1908).

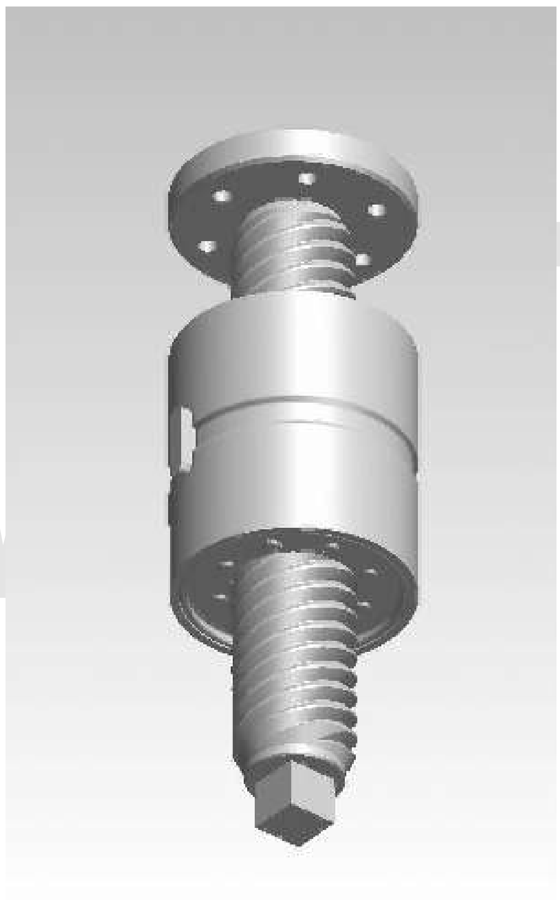
Rigid chain actuators were developed from “chain rammers” that used a single “ram chain” thrust from a magazine to load heavy-caliber ordinance into the breech of a cannon. Robert Matthews received a US patent for his “Mechanical Rammer” in 1901 which used a roller on the leading end of the chain to guide it and allow thrust without deflection. Developed more than a century ago, his rammer still bears a strong resemblance to many modern rigid chain actuators. In 1908 Oscar Knoch was awarded a US patent for his “Chain Rammer for Guns”. By orienting the folding side of the chain upward his ram chain acted as a self-supporting telescoping beam with negligible sag. Used in this manner the need for a separate guide was eliminated.

An early conception of chain used as a telescoping column instead a horizontal rammer was by Eldridge E. Long, who as awarded a US patent for his “Lifting Jack” in 1933, which he believed was “particularly adapted for use upon automobiles”. It used a double chain configuration, each chain linking solid bearing blocks that were stacked to resist compressive loads. In 1951, Yaichi Hayakawa was awarded a US patent for his “Interlocking Chain Stanchion” which eliminated bearing blocks by integrated the compressive path of force into the interlocking links of two roller-like chains. The zipper action of back-to-back interlocking chains provided guideless chain travel regardless of orientation and path of travel.

It should be noted that in 1941, prior to the double chain configuration, Karl Bender received a US patent for "Compression Resistant Chain" using three interlocking chains. In addition to the back-to-back arrangement of the typical interlocking chain actuator, a third chain was interlocked between the other two at a right angle. Perhaps due to their relative complexity, triple-chain actuators are not common.

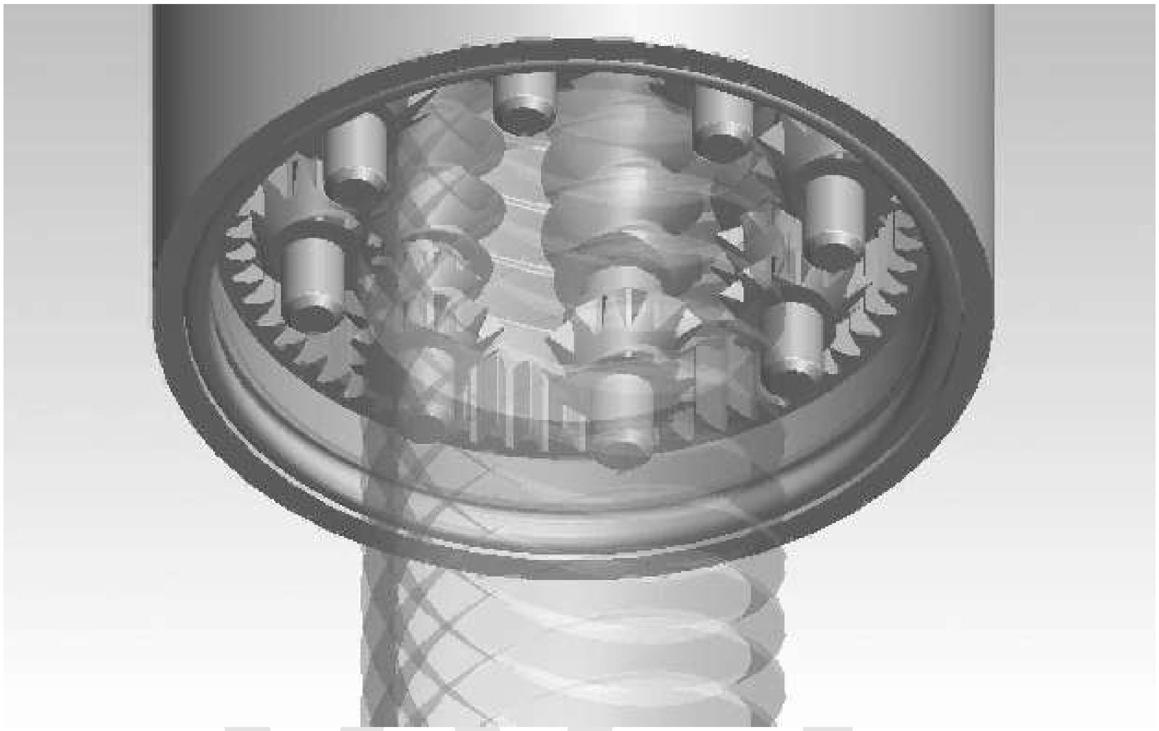
# Chapter- 15

# Roller Screw



A **roller screw**, also known as a **planetary roller screw** or **satellite roller screw**, is a low-friction precision mechanical device for converting rotational motion to linear motion, or vice versa. Planetary roller screws are most commonly used as the actuator mechanism in electro-mechanical linear actuators. Due to its complexity the roller screw is a relatively expensive actuator (as much as an order of magnitude more expensive than ball screws), but may be suitable for high-precision, high-speed, heavy-load, long-life and heavy-use applications.

## Principle of operation



Standard roller screw timing

A roller screw is a mechanical actuator similar to a ball screw that uses rollers as the load transfer elements between nut and screw instead of balls. The rollers are typically threaded but may also be grooved depending on roller screw type. Providing more bearing points than ball screws within a given volume, roller screws can be more compact for a given load capacity while providing similar efficiency (75%-90%) at low to moderate speeds, and maintain relatively high efficiency at high speeds. Roller screws can surpass ball screws in regard to positioning precision, load rating, rigidity, speed, acceleration, and lifetime. Standard roller screw actuators can achieve dynamic load ratings above 130 tons of force (exceeded in single-unit actuator capacity only by hydraulic cylinders).

The three main elements of a typical planetary roller screw are the screw shaft, nut and planetary roller. The screw, a shaft with a multi-start V-shaped thread, provides a helical raceway for multiple rollers radially arrayed around the screw and encapsulated by a threaded nut. The thread of the screw is typically identical to the internal thread of the nut. The rollers spin in contact with, and serve as low-friction transmission elements between screw and nut. The rollers typically have a single-start thread with convex flanks that limit friction at the rollers' contacts with screw and nut. The rollers typically orbit the screw as they spin (in the manner of planet gears to sun gear), and are thus known as planetary, or satellite, rollers. As with a lead screw or ball screw, rotation of the nut results in screw travel, and rotation of the screw results in nut travel.

For a given screw diameter and quantity of thread starts more rollers corresponds to higher static load capacity, but not necessarily to a higher dynamic load capacity. Preloaded split nuts and double nuts are available to eliminate backlash.

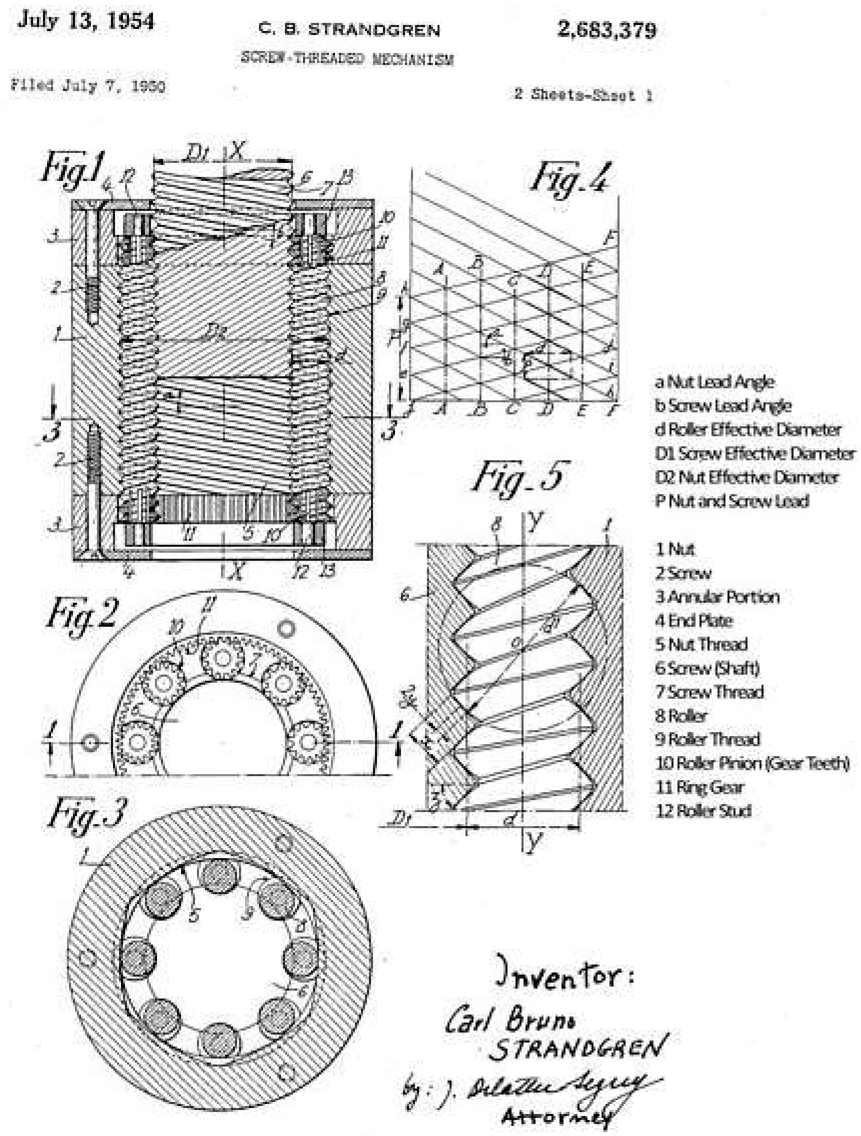
## Planetary roller screw types

Carl Bruno Strandgren developed some of the earliest effective forms of roller screws and was awarded US patents for such a "Screw-Threaded Mechanism" in 1954, and "Nut and Screw Devices" and the "Roller Screw" in 1965.

Roller screw types are defined by the motion of the rollers relative to the nut and screw. The four commercially available types of roller screw are *standard*, *inverted*, *recirculating*, and *bearing ring*.

*Differential roller screws*, typically variants of the standard and recirculating types, are also commercially available. Differential roller screws modify the rotational speed ratios between the rollers and the screw by varying the flank angles and contact points of the threads or grooves. In that way differential roller screws change the effective lead of the screw. William J. Roantree received a US patent for the "Differential Roller Nut" in 1968.

## Standard planetary roller screw




Patent drawing for standard roller screw (1954), with legend.

The standard planetary roller screw is also known as the non-recirculating roller screw. The lack of axial movement of the roller relative to the nut, and the gearing of rollers to nut, are definitive of the standard type of roller screw.

The nut and screw have identical multiple-start threads. The rollers have a single-start thread with an angle matching the nut thread. The matched thread angle prevents axial movement between the nut and the roller as the rollers spin. The nut assembly includes spacer rings and ring gears that position and guide the rollers. The spacer rings, which rotate within the ring gears, have equidistant holes that act as rotary bearings for the smooth pivot ends (studs) of the rollers. The ring gears time the spinning and orbit of the rollers about the screw axis by engaging gear teeth near the ends of the rollers. The spacer rings rotate on axis with the screw in unison with the orbit of the rollers. The spacer rings float relative to the nut, axially secured by retaining rings, because they spin around the screw at a lower frequency (angular velocity) than the nut.

## Configuration

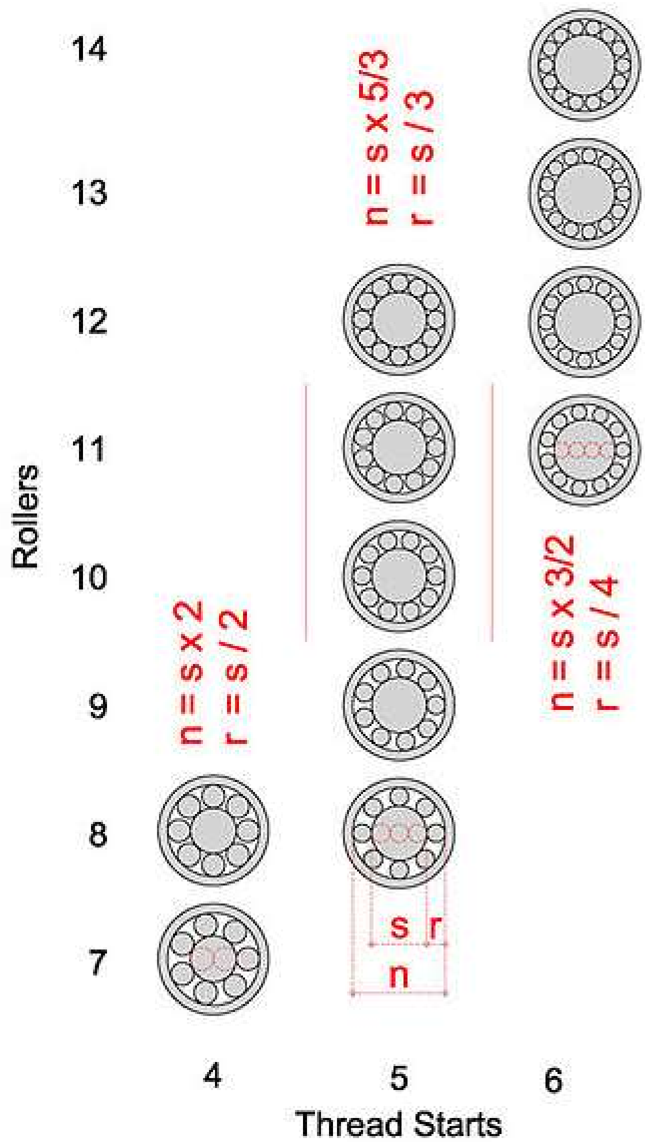
Standard roller screws are typically identified by screw diameter (typically ranging from 3.5mm – 200mm) and lead (1mm – 42mm). The threading of the screw (3 – 6 starts) is either rolled (lower capacity) or ground (higher capacity). The diameters of the nut and rollers (7 – 14 in quantity) are simple functions of the screw diameter and lead.

Where:

| | |
|---|---|
| $s_d$ = effective screw diameter * | $t$ = thread starts on nut and screw |
| $r_d$ = effective roller diameter | $l$ = screw lead |
| $n_d$ = effective nut inside diameter | $p$ = roller thread pitch |

* effective diameter is also known as pitch diameter

The following relationships apply to standard and inverted roller screws:




Common configurations of standard roller screws

$$n_d = \frac{s_d \cdot t}{t-2} \quad \therefore$$

nut to screw gear ratio

$(n_d : s_d) = 1 : (t / (t - 2))$

$$r_d = \frac{s_d}{t-2} \quad \therefore$$

roller to screw gear ratio

$(r_d : s_d) = 1 : (t - 2)$

$$r_d = \frac{n_d}{t}$$ $\therefore$ roller to nut gear ratio $(r_d : n_d) = 1 : t$

$$r_d = \frac{n_d - s_d}{2}$$

$$p = \frac{l}{t}$$ $\therefore$ ratio of roller thread pitch to screw lead $(p : l) = 1 : t$

For example, if

Screw: 30mm diameter, 20mm lead, 5 start thread

then

Rollers: 10mm diameter rollers, 4mm thread pitch
Nut: 50mm effective diameter.

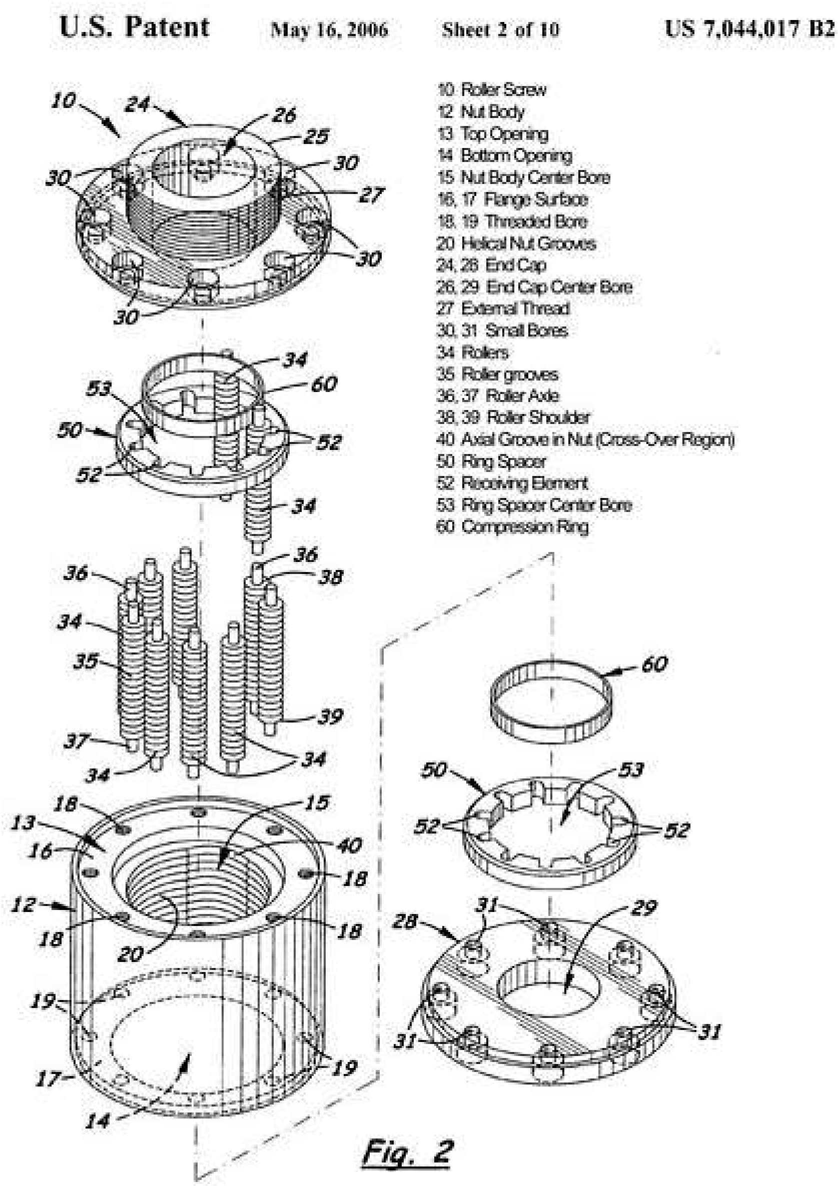
## Inverted roller screw

The inverted planetary roller screw is also known as the reverse roller screw. The lack of axial movement of the roller relative to the screw, and the gearing of rollers to screw, are definitive of the inverted type of planetary roller screw. This type of roller screw was developed simultaneously with the standard roller screw.

Inverted roller screws operate on the same principles of standard roller screws except that the function of the nut and screw is reversed in relation to the rollers. The rollers move axially within the nut, which is elongated to accommodate the full extent of screw shaft travel. The threaded portion of the screw shaft is limited to the threaded length of the rollers. The non-threaded portion of the screw shaft can be a smooth or non-cylindrical shape. The ring gear is replaced by gear teeth above and below the threaded portion of the screw shaft.

Aside from the inversion of the relationship of rollers to nut and screw, the configuration and relationships of inverted roller screws match those of standard roller screws.

## Recirculating roller screw



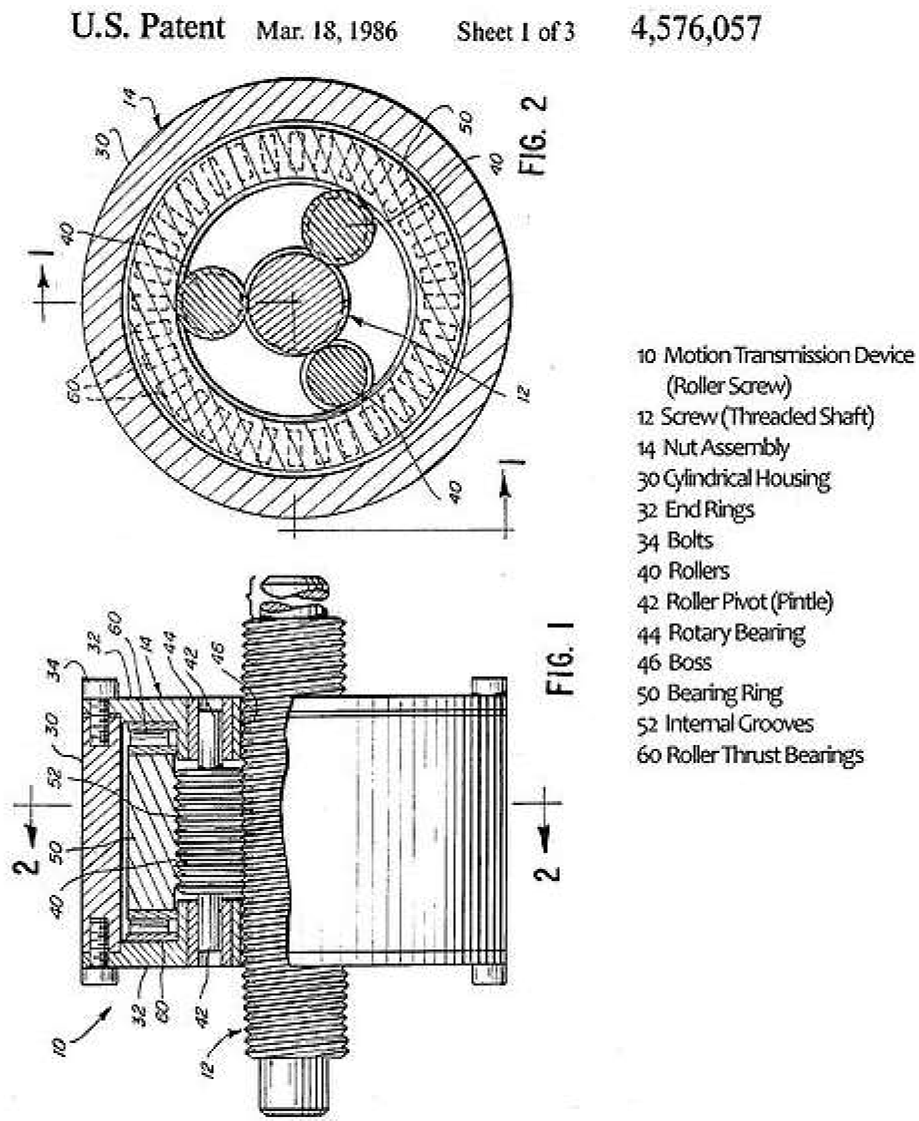

Cage-less recirculating roller screw patent drawing (2006), with legend.

The recirculating type of planetary roller screw is also known as a recycling roller screw. A recirculating roller screw can provide a very high degree of positional accuracy by using minimal thread leads. The rollers of a recirculating roller screw move axially within the nut until being reset after one orbit about the screw. Recirculating roller screws do not employ ring gears. Carl Bruno Strandgren was awarded a US Patent for the recirculating roller screw in 1965.

The screw and nut may have very fine identical single- or two-start threads. Recirculating rollers are grooved (instead of threaded) so they move axially during spinning engagement with the threads of the nut and screw, shifting up or down by one lead of thread after completing an orbit around the screw. The nut assembly typically includes a slotted cage and cam rings. The cage captivates the rollers in elongated slots, equally spacing the rollers while permitting rotation and axial motion of the rollers. The cam rings have opposing cams aligned with an axial groove in the wall of the nut. After a roller completes an orbit about the nut it is released into the groove, disengages from nut and screw, and is pushed between the cams to the axial midpoint of the nut assembly (shifting by a distance equal to the lead of the screw). Returned to its starting position, and reengaged to nut and screw, the roller may then orbit the screw once again.

In 2006, Charles C. Cornelius and Shawn P. Lawlor received a patent for a cage-less recirculating roller screw system. As with the traditional recirculating roller screw system, rollers disengage from the screw when they come upon an axial groove in the wall of the nut. The system differs in that the rollers are continually engaged by the nut, and the axial groove of the nut is threaded. Non-helical threads in the axial groove of the nut return the roller to its axial starting position (after completion of an orbit). Non-circular compression rings, or cam rings, at opposite ends of the rollers (roller axles) apply constant pressure between rollers and nut, synchronizing roller rotation and thrusting the rollers into the nut's axial groove. Lacking ring gears and roller cage, cage-less recirculating roller screws can be relatively efficient and, as a result, permit higher dynamic capacities for some screw shaft diameters.

## Bearing ring roller screw




Patent Drawing for Spiracon Roller Screw (1986), with legend.

In 1986 Oliver Saari was awarded a patent for a bearing ring roller screw, commonly referred to by its trademark, Spiracon. This type matches the orbit of the rollers to the rotation of the nut assembly. The actuator contains more load transfer elements than the other types, a bearing ring and thrust bearings, but manufacture of component parts is relatively simple (e.g. gearing teeth may be eliminated).

In the other roller screw types above, loads are transferred from the nut through the rollers to the screw (or in the reverse order). In this type of actuator, thrust bearings and a freely-rotating internally-grooved bearing ring transfer loads between the rollers and the nut.

The screw has a multi-start thread. The rollers and encapsulating rotating ring are identically grooved, not threaded, so there is no axial movement between the two. The nut assembly includes a cylindrical housing capped by non-rotating spacer rings. The spacer rings have equidistant holes that act as rotary bearings for the smooth pivot ends (studs) of the rollers. Roller-type thrust bearings between the spacer rings and bearing ring permit free rotation of the bearing ring while transferring the axial load between the two.

The rollers act as the “threads” of the nut assembly, causing axial movement of the rotating screw due to their orbital restraint. Screw rotation spins the rollers, which spin the bearing ring, dissipating the load-induced friction along the way.

Timothy A. Erhart was awarded a US patent in 1996 for a linear actuator effectively incorporating an inverted bearing ring roller screw. The screw shaft is grooved the length of and to match the grooved rollers, which travel with the shaft. The bearing ring is elongated and internally threaded for the length of screw shaft travel. The nut assembly housing and sealed end ring forms the exterior of the actuator assembly.