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Editorial

The rationale for federal support for basic research is well established, but the best policy for implementing this principle remains open to debate.The Nobel Prize–winning economist Robert E. Lucas Jr. wrote that once one starts thinking about long-run growth and economic development, "it is hard to think about anything else." Although I don't think I would go quite that far, it is certainly true that relatively small diferences in rates of economic growth, maintained over a sustained period, can have enormous implications for material living standards. A growth rate of output per person of 2.5% per year doubles average living standards in 28 years—about one generation—whereas output per person growing at what seems a modestly slower rate of 1.5% a year leads to a doubling in average living standards in about 47 years—roughly two generations. Compound interest is powerful! Of course, factors other than aggregate economic growth contribute to changes in living standards for different segments of the population, including shifts in relative wages and in rates of labor market participation. Nonetheless, if output per person increases more rapidly, the prospects for greater and more broad-based prosperity are significantly enhanced. Economic policy affects innovation and long-run economic growth in many ways. A stable macroeconomic environment; sound public finances; and well-functioning financial, labor, and product markets all support innovation, entrepreneurship, and growth, as do effective tax, trade, and regulatory policies. Policies directed at objectives such as the protection of intellectual property rights and the promotion of research and development, or R&D, promote innovation and technological change more directly.I will focus on one important component of innovation policy: government support for R&D. As I have already suggested, the effective commercial application of new ideas involves much more than just pure research. Many other factors are relevant, including the extent of market competition, the intellectual property regime, and the availability of financing for innovative enterprises. That said, the tendency of the market to supply too little of certain types of R&D provides a rationale for government intervention; and no matter how good the policy environment, big new ideas are often ultimately rooted in well-executed R&D.

Governments in many countries directly support scientific and technical research; for example, through grant-providing agencies or through tax incentives (like the R&D tax credit). In addition, the governments of the United States and many other countries run their own research facilities, including facilities focused on nonmilitary applications such as health. The primary economic rationale for a government role in R&D is that, without such intervention, the private market would not adequately supply certain types of research. The argument, which applies particularly strongly to basic or fundamental research, is that the full economic value of a scientific advance is unlikely to accrue to its discoverer, especially if the new knowledge can be replicated or disseminated at low cost. For example, James Watson and Francis Crick received a minute fraction of the economic benefits that have flowed from their discovery of the structure of DNA. If many people are able to exploit, or otherwise benefit from, research done by others, then the total or social return to research may be higher on average than the private return to those who bear the costs and risks of innovation. As a result, market forces will lead to underinvestment in R&D from society's perspective, providing a rationale for government intervention.

—Editors

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Drugs and their Side Effects

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In medicine, a side effect is an effect, whether therapeutic or adverse, that is secondary to the one intended; although the term is predominantly employed to describe adverse effects, it can also apply to beneficial, but unintended, consequences of the use of a drug.

Occasionally, drugs are prescribed or procedures performed specifically for their side effects; in that case, said side effect ceases to be a side effect, and is now an intended effect. For instance, X-rays were historically (and are currently) used as an imaging technique; the discovery of their oncolytic capability led to their employ in radiotherapy (ablation of malignant tumours).

If it results from an unsuitable or incorrect dosage or procedure, this is called a medical error and not a complication. Adverse effects are sometimes referred to as "iatrogenic" because they are generated by a physician/treatment. Some adverse effects only occur only when starting, increasing or discontinuing a treatment.

Using a drug or other medical intervention which is contraindicated may increase the risk of adverse effects. Adverse effects may cause complications of a disease or procedure and negatively affect its prognosis. They may also lead to non-compliance with a treatment regimen.

The harmful outcome is usually indicated by some result such as morbidity, mortality, alteration in body weight, levels of enzymes, loss of function, or as a pathological change detected at the microscopic, macroscopic or physiological level. It may also be indicated by symptoms reported by a patient. Adverse effects may cause a reversible or irreversible change, including an increase or decrease in the susceptibility of the individual to other chemicals, foods, or procedures, such as drug interactions.

Classification

In terms of drugs, adverse events may be defined as: "Any untoward medical occurrence in a patient or clinical investigation subject administered a pharmaceutical product and which does not necessarily have to have a causal relationship with this treatment."

In clinical trials, a distinction is made between adverse events and serious adverse events. Generally, any event which causes death, permanent damage, birth defects, or requires hospitalization is considered an SAE. The results of these trials are often included in the labelling of the medication to provide information both for patients and the prescribing physicians.

The term "life-threatening" in the definition of "serious" refers to an event in which the patient was at risk of death at the time of the event; it does not refer to an event which hypothetically might have caused death if it were more severe.

Reporting Systems

In many countries, adverse effects are required by law to be reported, researched in clinical trials and included into the patient information accompanying medical devices and drugs for sale to the public. Investigators in human clinical trials are obligated to report these events in clinical study reports. Research suggests that these events are often inadequately reported in publicly available reports. Because of the lack of these data and uncertainty about methods for synthesising them, individuals conducting systematic reviews and meta-analyses of therapeutic interventions often unknowingly overemphasise health benefit. To balance the overemphasis on benefit, scholars have called for more complete reporting of harm from clinical trials.

United Kingdom

The Yellow Card Scheme is a United Kingdom initiative run by the Medicines and Healthcare products Regulatory Agency (MHRA) and the Commission on Human Medicines (CHM) to gather information on adverse effects to medicines. This includes all licensed medicines, from medicines issued on prescription to medicines bought over the counter from a supermarket. The scheme also includes all herbal supplements and unlicensed medicines found in cosmetic treatments. Adverse drug reactions (ADRs) can be reported by a number of health care professionals including physicians, pharmacists and nurses, as well as patients.

United States

In the United States several reporting systems have been built, such as the Vaccine Adverse Event Reporting System (VAERS), the Manufacturer and User Facility Device Experience Database (MAUDE) and the Special Nutritionals Adverse Event Monitoring System. MedWatch is the main reporting center, operated by the Food and Drug Administration.

Australia

In Australia, adverse effect reporting is administered by the Adverse Drug Reactions Advisory Committee (ADRAC), a subcommittee of the Australian Drug Evaluation Committee (ADEC). Reporting is voluntary, and ADRAC requests healthcare professionals to report all adverse reactions to its current drugs of interest, and serious adverse reactions to any drug. ADRAC publishes the Australian Adverse Drug Reactions Bulletin every two months.The Government's Quality Use of Medicines programme is tasked with acting on this reporting to reduce and minimize the number of preventable adverse effects each year.

New Zealand

Adverse reaction reporting is an important component of New Zealand's pharmacovigilance activities. The Centre for Adverse Reactions Monitoring (CARM) in Dunedin is New Zealand's national monitoring centre for adverse reactions. It collects and evaluates spontaneous reports of adverse reactions to medicines, vaccines, herbal products and dietary supplements from health professionals in New Zealand. Currently the CARM database holds over 80,000 reports and provides New Zealand-specific information on adverse reactions to these products, and serves to support clinical decision making when unusual symptoms are thought to be therapy related

Canada

In Canada, adverse reaction reporting is an important component of the surveillance of marketed health products conducted by the Health Products and Food Branch (HPFB) of Health Canada. Within HPFB, the Marketed Health Products Directorate leads the coordination and implementation of consistent monitoring practices with regards to assessment of signals and safety trends, and risk communications concerning regulated marketed health products.

MHPD also works closely with international organizations to facilitate the sharing of information. Adverse reaction reporting is mandatory for the industry and voluntary for consumers and health professionals.

Limitations

In principle, medical professionals are required to report all adverse effects related to a specific form of therapy. In practice, it is at the discretion of the professional to determine whether a medical event is at all related to the therapy. For example, a leg fracture in a skiing accident in a patient who years before took antibiotics for pneumonia is not likely to get reported.

As a result, routine adverse effects reporting often may not include long-term and subtle effects that may ultimately be attributed to a therapy.

Part of the difficulty is identifying the source of a complaint. A headache in a patient taking medication for influenza may be caused by the underlying disease or may be an adverse effect of the treatment. In patients with end-stage cancer, death is a very likely outcome and whether the drug is the cause or a bystander is often difficult to discern.

Adverse Effects by Situation

Adverse Effects of Medical Procedures

Surgery may have a number of undesirable or harmful effects, such as infection, hemorrhage, inflammation, scarring, loss of function, or changes in local blood flow. They can be reversible or irreversible, and a compromise must be found by the physician and the patient between the beneficial or life-saving consequences of surgery versus its adverse effects. For example, a limb may be lost to amputation in case of untreatable gangrene, but the patient's life is saved. Presently, one of the greatest advantages of minimally invasive surgery, such as laparoscopic surgery, is the reduction of adverse effects.

Other nonsurgical physical procedures, such as high-intensity radiation therapy, may cause burns and alterations in the skin. In general, these therapies try to avoid damage to healthy tissues while maximizing the therapeutic effect.

Vaccination may have adverse effects due to the nature of its biological preparation, sometimes using attenuated pathogens and toxins. Common adverse effects may be fever, malaise and local reactions in the vaccination site. Very rarely, there is a serious adverse effect, such as eczema vaccinatum, a severe, sometimes fatal complication which may result in persons who have eczema or atopic dermatitis.

Diagnostic procedures may also have adverse effects, depending much on whether they are invasive, minimally invasive or noninvasive. For example, allergic reactions to radiocontrast materials often occur, and a colonoscopy may cause the perforation of the intestinal wall.

Adverse Effects of Drugs

Adverse effects can occur as a collateral or side effect of many interventions, but they are particularly important in pharmacology, due to its wider, and sometimes uncontrollable, use by way of self-medication. Thus, responsible drug use becomes an important issue here. Adverse effects, like therapeutic effects of drugs, are a function of dosage or drug levels at the target organs, so they may be avoided or decreased by means of careful and precise pharmacokinetics, the change of drug levels in the organism in function of time after administration.

Adverse effects may also be caused by drug interaction. This often occurs when patients fail to inform their physician and pharmacist of all the medications they are taking, including herbal and dietary supplements. The new medication may interact agonistically or antagonistically (potentiate or decrease the intended therapeutic effect), causing significant morbidity and mortality around the world. Drug-drug and food-drug interactions may occur, and so-called "natural drugs" used in alternative medicine can have dangerous adverse effects. For example, extracts of St John's wort (*Hypericum perforatum*), a phytotherapic used for treating mild depression are known to cause an increase in the cytochrome P450 enzymes responsible for the metabolism and elimination of many drugs, so patients taking it are likely to experience a reduction in blood levels of drugs they are taking for other purposes, such as cancer chemotherapeutic drugs, protease inhibitors for HIV and hormonal contraceptives.

The scientific field of activity associated with drug safety is increasingly government-regulated, and is of major concern for the public, as well as to drug manufacturers. The distinction between adverse and nonadverse effects is a major undertaking when a new drug is developed and tested before marketing it. This is done in toxicity studies to determine the nonadverse effect level (NOAEL). These studies are used to define the dosage to be used in human testing (phase I), as well as to calculate the maximum admissible daily intake. Imperfections in clinical trials, such as insufficient number of patients or short duration, sometimes lead to public health disasters, such as those of fenfluramine (the so-called fen-phen episode), thalidomide and, more recently, of cerivastatin (Baycol, Lipobay) and rofecoxib (Vioxx), where drastic adverse effects were observed, such as teratogenesis, pulmonary hypertension, stroke, heart disease, neuropathy, and a significant number of deaths, causing the forced or voluntary withdrawal of the drug from the market.

Most drugs have a large list of nonsevere or mild adverse effects which do not rule out continued usage. These effects, which have a widely variable incidence according to individual sensitivity, include nausea, dizziness, diarrhea, malaise, vomiting, headache, dermatitis, dry mouth, etc. These can be considered a form of pseudo-allergic reaction, as not all users experience these effects; many users experience none at all.

Drugs contain side effects which is the reason why commercials or advertisements put many disclaimers about the unwanted symptoms after taking the drug(s).

Examples of adverse effects associated with specific medications

- Abortion, miscarriage or uterine hemorrhage associated with misoprostol (Cytotec), a labor-inducing drug (this is a case where the adverse effect has been used legally and illegally for performing abortions)
- Addiction to many sedatives and analgesics, such as diazepam, morphine, etc.
- Birth defects associated with thalidomide
- Bleeding of the intestine associated with aspirin therapy
- Cardiovascular disease associated with COX-2 inhibitors (i.e. Vioxx)
- Deafness and kidney failure associated with gentamicin (an antibiotic)
- Death, following sedation, in children using propofol (Diprivan)
- Depression or hepatic injury caused by interferon
- Diabetes caused by atypical antipsychotic medications (neuroleptic psychiatric drugs)
- Diarrhea caused by the use of orlistat (Xenical)
- Erectile dysfunction associated with many drugs, such as antidepressants
- Fever associated with vaccination
- Glaucoma associated with corticosteroid-based eye drops
- Hair loss and anemia may be caused by chemotherapy against cancer, leukemia, etc.
- Headache following spinal anaesthesia
- Hypertension in ephedrine users, which prompted FDA to remove the dietary supplement status of ephedra extracts
- Insomnia caused by stimulants, methylphenidate (Ritalin), Adderall, etc.
- Lactic acidosis associated with the use of stavudine (Zerit, for HIV therapy) or metformin (for diabetes)
- Mania caused by corticosteroids
- Liver damage from paracetamol

- Melasma and thrombosis associated with use of estrogen-containing hormonal contraception, such as the combined oral contraceptive pill
- Priapism associaded with the use of sildenafil
- Rhabdomyolysis associated with statins (anticholesterol drugs)
- Seizures caused by withdrawal from benzodiazepines
- Drowsiness or increase in appetite due to antihistamine use. Some antihistamines are used in sleep aids explicitly because they cause drowsiness.
- Stroke or heart attack associated with sildenafil (Viagra), when used with nitroglycerin
- Suicide, increased tendency associated to the use of fluoxetine and other selective serotonin reuptake inhibitor (SSRI) antidepressants
- Tardive dyskinesia associated with long-term use of metoclopramide and many antipsychotic medications

Controversies

Sometimes, putative medical adverse effects are regarded as controversial and generate heated discussions in society and lawsuits against drug manufacturers. One example is the recent controversy as to whether autism was linked to the MMR vaccine (or by thiomersal, a mercury-based preservative used in some vaccines). No link has been found in several large studies, and despite removal of thimerosal from vaccines a decade ago the rate of autism has not decreased as would be expected if it had been the causative agent.

Another instance is the potential adverse effects of silicone breast implants, which led to hundreds of thousands of litigations against manufacturers of gel-based implants, due to allegations of damage to the immune system which have not yet been conclusively proven.

Due to the exceedingly high impact on public health of widely used medications, such as hormonal contraception and hormone replacement therapy, which may affect millions of users, even marginal probabilities of adverse effects of a severe nature, such as breast cancer, have led to public outcry and changes in medical therapy, although its benefits largely surpassed the statistical risks.

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Mechatronics

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Mechatronics is a design process that includes a combination of mechanical engineering, electrical engineering, telecommunications engineering, control engineering and computer engineering. Mechatronics is a multidisciplinary field of engineering, that is to say, it rejects splitting engineering into separate disciplines. Originally, mechatronics just included the combination of mechanics and electronics, hence the word is a combination of mechanics and electronics; however, as technical systems have become more and more complex the word has been broadened to include more technical areas.

The word "Mechatronics" was originally "Japanese-English" word created by Mr. Tetsuro Mori, who was an engineer of Yaskawa Electric Corporation and the word "Mechatronics" was registered as Trademark by the company in Japan with the registration number of "46-32714" in 1971. However afterward, the company released the right of using the word to public, and the word "Mechatronics" didn't only stay in Japan, but also introduced as a native English word. Nowadays, the word is translated in each language and the word is considered as an essential term for industry.

French standard NF E 01-010 gives the following definition: "approach aiming at the synergistic integration of mechanics, electronics, control theory, and computer science within product design and manufacturing, in order to improve and/or optimize its functionality".

Many people treat "mechatronics" as a modern buzzword synonymous with "electromechanical engineering". However, other people draw a distinction between an "electromechanical component"—does not include a computer; an electro-mechanical computer (such as the Z4)—does not include an electronic computer; vs. a "mechatronic system"—a computer-controlled mechanical system, including both an electronic computer and electromechanical components.

Description

A mechatronics engineer unites the principles of mechanics, electronics, and computing to generate a simpler, more economical and reliable system. The term "mechatronics" was coined by Tetsuro Mori, the senior engineer of the Japanese company Yaskawa in 1969. An industrial robot is a prime example of a mechatronics system; it includes aspects of electronics, mechanics, and computing to do its day-to-day jobs.

Engineering cybernetics deals with the question of control engineering of mechatronic systems. It is used to control or regulate such a system. Through collaboration, the mechatronic modules perform the production goals and inherit flexible and agile manufacturing properties in the production scheme. Modern production equipment consists of mechatronic modules that are integrated according to a control architecture. The most known architectures involve hierarchy, polyarchy, heterarchy, and hybrid. The methods for achieving a technical effect are described by control algorithms, which might or might not utilize formal methods in their design. Hybrid systems important to mechatronics include production systems, synergy drives, planetary exploration rovers, automotive subsystems such as anti-lock braking systems and spin-assist, and everyday equipment such as autofocus cameras, video, hard disks, and CD players.

Physical Implementations

Mechanical modelling calls for modelling and simulating physical complex phenomenon in the scope of a multi-scale and multi-physical approach. This implies to implement and to manage modelling and optimization methods and tools, which are integrated in a systemic approach. The specialty is aimed at students in mechanics who want to open their mind to systems engineering, and able to integrate different physics or technologies, as well as students in mechatronics who want to increase their knowledge in optimization and multidisciplinary simulation technics. The specialty educates students in robust and/or optimized conception methods for structures or many technological systems, and to the main modelling and simulation tools used in R&D. Special courses are also proposed for original applications (multi-materials composites, innovating transducers and actuators, integrated systems, …) to prepare the students to the coming breakthrough in the domains covering the materials and the systems. For some mechatronic systems, the main issue is no longer how to implement a control system, but how to implement actuators. Within the mechatronic field, mainly two technologies are used to produce movement/motion.

Variant of the Field

An emerging variant of this field is biomechatronics, whose purpose is to integrate mechanical parts with a human being, usually in the form of removable gadgets such as an exoskeleton. Such an entity is often identified in science fiction as a cyborg. This is the "real-life" version of cyberware.

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Analysis of Hierarchies for Object Recognition Techniques

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Every day we recognize a multitude of familiar and novel objects. We do this with little effort, despite the fact that these objects may vary somewhat in form, colour, texture, etc. Objects are recognized from many different vantage points (from the front, side, or back), in many different places, and in different sizes. Objects can even be recognized when they are partially obstructed from view.

While it may be obvious that people are capable of recognizing objects under many variations in conditions, it has been thought that pigeons may not possess the same range of capabilities. It has been proposed that pigeons act as "perceptrons," by analyzing simple features of objects and using those features to recognize objects. If the pigeon were a perceptron, then it would not be able to recognize an object that varied slightly in form or was seen from a novel viewpoint because the features would be altered. Moreover, a pigeon would be unable to discriminate between two objects that contained the same features, but with a different organization.

Object recognition is the ability to perceive an object's physical properties (such as shape, colour and texture) and apply semantic attributes to the object, which includes the understanding of its use, previous experience with the object and how it relates to others.

One model of object recognition, based on neuropsychological evidence, provides information that allows us to divide the process into four different stages.

Stage 1 Processing of basic object components, such as colour, depth, and form.

Stage 2 These basic components are then grouped on the basis of similarity, providing information on distinct edges to the visual form. Subsequently, figure-ground segregation is able to take place.

Stage 3 The visual representation is matched with structural descriptions in memory.

Stage 4 Semantic attributes are applied to the visual representation, providing meaning, and thereby recognition.

It should be noted that, within these stages, there are more specific processes that take place to complete the different processing components. In addition, other existing models propose integrative hierarchies (top-down and bottom-up), as well as parallel processing, as opposed to this general bottom-up hierarchy.

Visual recognition processing has been typically viewed as a bottom-up hierarchy in which information is processed sequentially with increasing complexities, where lower-level cortical processors, such as the primary visual cortex, are at the bottom of the processing hierarchy and higher-level cortical processors, such as the inferotemporal cortex (IT), are at the top, where recognition is facilitated. A most recognized bottom-up hierarchical theory is David Marr's theory of vision. In contrast, an increasingly popular recognition processing theory, is that of top-down processing. One model, proposed by Moshe Bar (2003), describes a "shortcut" method in which early visual inputs are sent, partially analyzed, from the early visual cortex to the prefrontal cortex (PFC). Possible interpretations of the crude visual input is generated in the PFC and then sent to the inferotemporal cortex (IT) subsequently activating relevant object representations which are then incorporated into the slower, bottom-up process. This "shortcut" is meant to minimize the amount of object representations required for matching thereby facilitating object recognition. Lesion studies have supported this proposal with findings of slower response times for individuals with PFC lesions, suggesting use of only the bottom-up processing.

A significant aspect of object recognition is that of object constancy: the ability to recognize an object across varying viewing conditions. These varying conditions include object orientation, lighting, and object variability (size, colour, and other within-category differences). For the visual system to achieve object constancy, it must be able to extract a commonality in the object description across different viewpoints and the retinal descriptions.

The Object Recognition module provides a way to identify specific trained objects within the current image. Once the module is trained with sample template images it will identify those objects within the current image depending on the filtered parameters of confidence, size, rotation, etc.

Serveral of the techniques will account for different object sizes, location and in plane rotation (roll) of the object aswell as variations in lighting and contrast. It will NOT account for significant rotation of the object in the X and Y (pan and tilt) directions. Should you need to identify a 3D object in any orientation you will need to include template examples of each orientation.

Templates are created by including images into a folder that is used as the training samples for this module. Thus you can use any image editing application to edit and manage those templates as needed. Note, it is always best just to include the object to be identified without any background parts of the image. Only one folder at a time can be selected into a single object Recognition module. By changing the folder you change those objects that are to be identified. Note that you can also use more than one Object Recognition module within the pipeline.

It is recommended to use as large a template image as possible that will appear in the scene. One that encompasses as much of the image size as possible is best. This is because the Object Recognition module will only seek out objects from the template size down to a minimum of 1/3th the template size. Thus having the template contain the most amount of detail will provide the best results. If you specify too large an object then smaller versions of the objects may not be recognized as they may fall below the 30% size limit.

Several recognition methods are provided. As many objects/ environments differ in task the module provides various methods that can be switched between in order to determine the best technique for your use. Note that while you can switch between the techniques by selecting the appropriate radio button the module will NOT update the template database when you switch to that method. Thus if you add a new image and want to experiment with all techniques you will have to switch to each method and you MUST press the Train button in order to ensure that the template database is up-to-date.

Feature Points

This method will identify interesting points within the template using a modified fast Harris feature detector and match those points with those detected within the current image. The identified points are typically corner-like points that exhibit restraining forces in each direction (i.e. the highest edge signal in both X and Y directions will be maximal at the point's position). This helps to stabilize point choices in both the template and image such that most (but not all) points will be detected between the template and image. Once this correlation has been done the most likely template is then tested for at that location using a slower cross correlation technique. This technique is a good standard technique assuming there is enough internal texture within the object (think of a book cover) and is fast enough for most purposes.

Keep in mind that as this technique mainly uses corners as features motion blur will cause all those features to disappear and therefore not match correctly. Thus if you have a lot of motion blur within images you wish to match against you may need to decide on an alternative technique.

Objects to be detected MUST have internal texture as this module relies in inner feature points as the primary object identification technique. If you are interested in just the shape of an object the Shape Matching module will be of better use or the Shape method mentioned below.

This method will check for translation, scale and orientation.

Shape

The shape matching method is similar to the Shape Matching module but this method works on intensity images contrary to the Shape Matching module which uses binary (Black&White) images. The shape method will analyze the template and current image for correspondences in shape and determine which parts of the image best contain a particular template.

Due to its reliance on shape the shape matching method can determine any size and orientation (including X and Y location) of the template image within the current image in about the same time as the cross correlation method does just X and Y location. Its speed and flexibility make the shape matching method very useful when limited internal template texture is available. For example, labels, street signs, text, logos, fiducials, etc. are all ideal templates for the shape matching method as they consist of non-textured areas and rely on shape as their primary form of identification.

It is recommended to use templates that are as large as possible with regards to what may be seen in the current image. If small templates are used and matching to larger possible targets in the current image is attempted a mismatch will occur as more detail will be in the current image than what is contained within the template. Thus it is important to use a template by cropping the largest size seen of the template in any test or production images.

This method will check for translation, scale and orientation.

Haar

A very popular face recognition technique uses Haar like filters to determine a set of pixels comparisons that best represents the concept of a face. While this has its uses the Haar technique can also be applied to specific object recognition. In this scenario many high intensity versus low intensity checks are created that when run in sequence will identify a template with a high probability. These checks can be run across the image very quickly and even adapt to size differences but they cannot be quickly rotated and thus are restricted to a single orientation.

Note that contrary to typical Haar training this method only requires ONE sample image to recognize the image and does NOT create a generic class based on that image. Thus if you want to recognize a class of objects you will need to include a couple images to best represent that class.

This method will check for translation and scale.

Cross Correlation

This method uses a well known and established technique for object recognition. The normalized cross correlation method has been widely used in many applications and can be one of the more stable techniques to use. The algorithm behind the cross correlation technique is, unfortunately, very CPU intensive and therefore should be used with care otherwise significant time may be spent on searching for objects within an image.

The cross correlation algorithm will use the template as saved in the trained folder and compare the template pixel by pixel at each pixel location within the current image. For example, if your template is 50x50 pixels large and your current image is 320x240 large this will mean there are approximately 2500x76800 = 192,000,000 pixel comparisons to make. This brute force method is entirely too slow for reasonable usage so tradeoffs between the size of the template versus the size of the image is made in order to speed up comparisons. While this does significantly speed up the algorithm some accuracy is forfeited as a result of that speed.

While normalized cross correlation does attempt to best deal with lighting changes between the template and the current image it is not performed within the image itself, thus while the template and current test image can be overall darker or lighter if a shadow is cast across the template that is not also within the current image the recognition process will fail.

To further reduce performance requirements the cross correlation method ONLY checks for templates of the same size AND orientation. Thus a template is only searched in the X and Y direction. If the current image contains the template at a different size it will NOT be recognized (but can still be tracked). If this is a requirement for your project and you decide to use cross correlation you will need to add the template in different sizes to the training folder. We recommend changing template sizes in increments of 10% which typically will provide adequate coverage.

If orientation is required you may attempt to use the Orient Image module prior to cross correlation which will help to specify a standard orientation prior to matching. Note that you will need to add the 180deg rotated template within the training folder if you decide on this direction as the Orient Image will most likely have a 180deg symmetry and therefore not always align to the same 180 direction.

For comparisons that are image to image (meaning NO change in X or Y) the cross correlation technique can be one of the fastest. Thus if you goal is to find an image that exists within a known database without any size, orientation, horizontal or vertical shifts the cross correlation technique is a good technique to attempt.

In data mining, hierarchical clustering is a method of cluster analysis which seeks to build a hierarchy of clusters.

In order to decide which clusters should be combined (for agglomerative), or where a cluster should be split (for divisive), a measure of dissimilarity between sets of observations is required. In most methods of hierarchical clustering, this is achieved by use of an appropriate metric (a measure of distance between pairs of observations), and a linkage criterion which specifies the dissimilarity of sets as a function of the pair-wise distances of observations in the sets.

Most object detection tasks in computer vision are computationally expensive because of a) the large amount of input data that has to be processed and b) the use of complex classifiers that are robust against pose and illumination changes. Speeding-up the classification is therefore of major concern when developing systems for real-world applications. In the following we investigate two methods for speed-ups: feature reduction and hierarchical classification.

There are basically two types of feature selection methods in the literature: filter and wrapper methods. Filter methods are preprocessing steps performed independently of the classification algorithm or its error criteria; PCA is an example of a filter method. Wrapper methods attempt to search through the space of feature subsets using the criterion of the classification algorithm to select the optimal feature subset. Wrapper methods can provide more accurate solutions than filter methods, but in general are more computationally expensive. We present a new wrapper method to reduce the dimensions of both input and feature space of an SVM. An alternative approach for speeding-up SVM classification has been proposed in by reducing the number of support vectors.

Feature reduction is a generic tool that can be applied to any classification problem. When dealing with a specific classification task we can use prior knowledge about the type of data to speed-up classification. Two assumptions hold for most vision-based object detection tasks: a) The vast majority of the analyzed patterns in an image belongs to the background class and b) most of the background patterns can be easily distinguished from the objects. Based on these two assumptions it is sensible to apply a hierarchy of classifiers. Fast classifiers remove large parts of the back-groundon the bottom and middle levels of the hierarchy and a more accurate but slower classifier performs the final detection on the top level. This idea falls into the framework of coarse-to-fine template matching and is also related to biologically motivated work on attention-based vision.

More recently a cascade of linear classifiers that have been trained using Ada Boost has been proposed in for frontal face detection. This idea is related to ours in the sense that it combines hierarchical classification with feature selection. However, in our approach the complexity of the classifiers in the hierarchy is not only controlled by the number of features (image resolution) but also by the class of decision functions (i.e. class of SVM kernel functions).

The bottom level of our hierarchy consists of a linear classifier that operates on low resolution patterns (9x9) while the top level consists of a non-linear classifier operating on higher resolution patterns (19x19).

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Electronic Design Automation

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Electronic design automation (EDA or ECAD) is a category of software tools for designing electronic systems such as printed circuit boards and integrated circuits. The tools work together in a design flow that chip designers use to design and analyze entire semiconductor chips.

Before EDA, integrated circuits were designed by hand, and manually laid out. Some advanced shops used geometric software to generate the tapes for the Gerber photoplotter, but even those copied digital recordings of mechanically drawn components. The process was fundamentally graphic, with the translation from electronics to graphics done manually. The best known company from this era was Calma, whose GDSII format survives.

By the mid-1970s, developers started to automate the design along with the drafting. The first placement and routing (Place and route) tools were developed. The proceedings of the Design Automation Conference cover much of this era.

The next era began about the time of the publication of "Introduction to VLSI Systems" by Carver Mead and Lynn Conway in 1980. This ground breaking text advocated chip design with programming languages that compiled to silicon. The immediate result was a considerable increase in the complexity of the chips that could be designed, with improved access to design verification tools that used logic simulation. Often the chips were easier to lay out and more likely to function correctly, since their designs could be simulated more thoroughly prior to construction. Although the languages and tools have evolved, this general approach of specifying the desired behaviour in a textual programming language and letting the tools derive the detailed physical design remains the basis of digital IC design today.

The earliest EDA tools were produced academically. One of the most famous was the "Berkeley VLSI Tools Tarball", a set of UNIX utilities used to design early VLSI systems. Still widely used is the Espresso heuristic logic minimizer and Magic.

Another crucial development was the formation of MOSIS, a consortium of universities and fabricators that developed an inexpensive way to train student chip designers by producing real integrated circuits.

The basic concept was to use reliable, low-cost, relatively low-technology IC processes, and pack a large number of projects per wafer, with just a few copies of each projects' chips. Cooperating fabricators either donated the processed wafers, or sold them at cost, seeing the programme as helpful to their own long-term growth.

Birth of Commercial EDA

1981 marks the beginning of EDA as an industry. For many years, the larger electronic companies, such as Hewlett Packard, Tektronix, and Intel, had pursued EDA internally. In 1981, managers and developers spun out of these companies to concentrate on EDA as a business. Daisy Systems, Mentor Graphics, and Valid Logic Systems were all founded around this time, and collectively referred to as DMV. Within a few years there were many companies specializing in EDA, each with a slightly different emphasis. The first trade show for EDA was held at the Design Automation Conference in 1984.

In 1981, the U.S. Department of Defense began funding of VHDL as a hardware description language. In 1986, Verilog, another popular high-level design language, was first introduced as a hardware description language by Gateway Design Automation. Simulators quickly followed these introductions, permitting direct simulation of chip designs: executable specifications. In a few more years, back-ends were developed to perform logic synthesis.

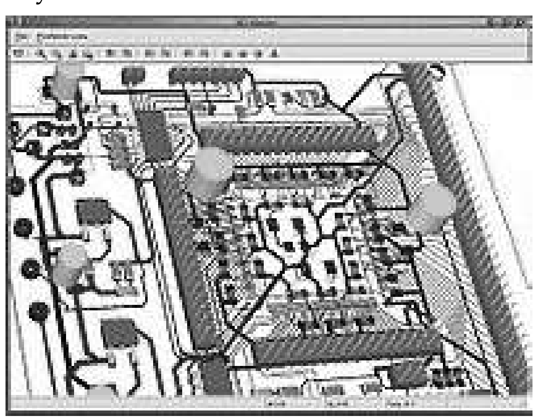

Figure: *3D PCB layout*

Current Status

Current digital flows are extremely modular. The front ends produce standardized design descriptions that compile into invocations of "cells,", without regard to the cell technology. Cells implement logic or other electronic functions using a particular integrated circuit technology.

Fabricators generally provide libraries of components for their production processes, with simulation models that fit standard simulation tools. Analog EDA tools are far less modular, since many more functions are required, they interact more strongly, and the components are (in general) less ideal.

EDA for electronics has rapidly increased in importance with the continuous scaling of semiconductor technology. Some users are foundry operators, who operate the semiconductor fabrication facilities, or "fabs", and design-service companies who use EDA software to evaluate an incoming design for manufacturing readiness. EDA tools are also used for programming design functionality into FPGAs.

Software Focuses

Design

- High-level synthesis (syn. behavioural synthesis, algorithmic synthesis) For digital chips
- Logic synthesis translation of abstract, logical language such as Verilog or VHDL into a discrete netlist of logic gates
- Schematic Capture For standard cell digital, analog, RF-like Capture CIS in Orcad by CADENCE and ISIS in Proteus
- Layout, usually schematic-driven layout, like Layout in Orcad by Cadence, ARES in Proteus

Simulation

- Transistor simulation – low-level transistor-simulation of a schematic/ layout's behaviour, accurate at device-level.
- Logic simulation – digital-simulation of an RTL or gate-netlist's digital (boolean 0/1) behaviour, accurate at boolean-level.
- Behavioural Simulation – high-level simulation of a design's architectural operation, accurate at cycle-level or interface-level.
- Hardware emulation – Use of special purpose hardware to emulate the logic of a proposed design. Can sometimes be plugged into a system in place of a yet-to-be-built chip; this is called in-circuit emulation.
- Technology CAD simulate and analyze the underlying process technology. Electrical properties of devices are derived directly from device physics.
- Electromagnetic field solvers, or just field solvers, solve Maxwell's equations directly for cases of interest in IC and PCB design. They are known for being slower but more accurate than the layout extraction above.

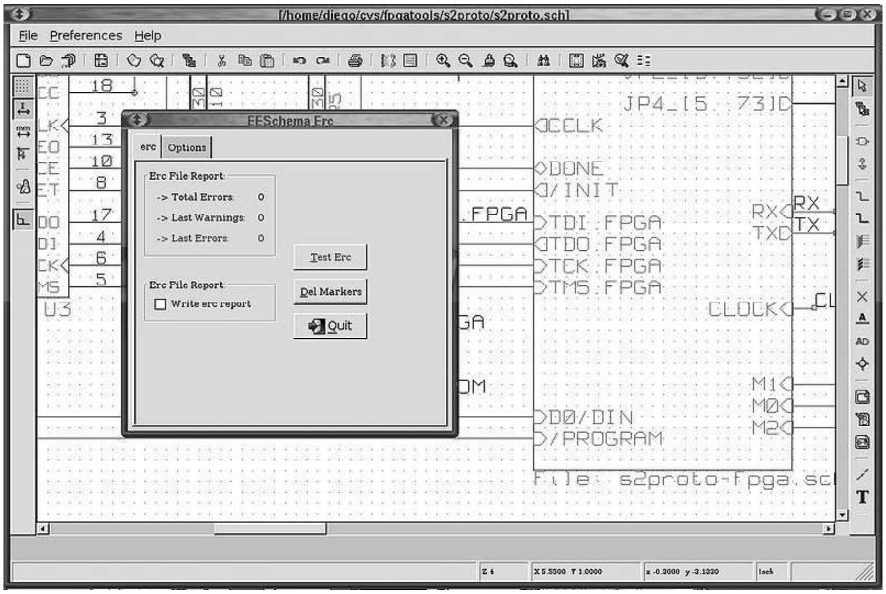

Figure: Schematic capture programme

Analysis and Verification

- Functional verification
- Clock Domain Crossing Verification (CDC check): Similar to linting, but these checks/tools specialize in detecting and reporting potential issues like data loss, meta-stability due to use of multiple clock domains in the design.
- Formal verification, also model checking: Attempts to prove, by mathematical methods, that the system has certain desired properties, and that certain undesired effects (such as deadlock) cannot occur.
- Equivalence checking: algorithmic comparison between a chip's RTL-description and synthesized gate-netlist, to ensure functional equivalence at the *logical* level.
- Static timing analysis: Analysis of the timing of a circuit in an input-independent manner, hence finding a worst case over all possible inputs.
- Physical verification, PV: checking if a design is physically manufacturable, and that the resulting chips will not have any function-preventing physical defects, and will meet original specifications.

Manufacturing Preparation

- Mask data preparation, MDP: generation of actual lithography photomask used to physically manufacture the chip.

- o Resolution enhancement techniques, RET – methods of increasing of quality of final photomask.
- o Optical proximity correction, OPC – up-front compensation for diffraction and interference effects occurring later when chip is manufactured using this mask.
- o Mask generation – generation of flat mask image from hierarchical design.
- o Automatic test pattern generation, ATPG – generates pattern-data to systematically exercise as many logic-gates, and other components, as possible.
- o Built-in self-test, or BIST – installs self-contained test-controllers to automatically test a logic (or memory) structure in the design

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GPS Navigation Device

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The Global Positioning System (GPS) is a satellite-based navigation system made up of a network of 24 satellites placed into orbit by the U.S. Department of Defense. Military actions was the original intent for GPS, however in the 1980s, the U.S. government decided to allow the GPS programme to be used by civilians. Weather conditions do not affect the ability for GPS to work. The systems works 24/7 anywhere in the world. There are no subscription fees or setup charges to use GPS.

GPS devices may have capabilities such as:

- maps, including streets maps, displayed in human readable format via text or in a graphical format,
- turn-by-turn navigation directions to a human in charge of a vehicle or vessel via text or speech,
- directions fed directly to an autonomous vehicle such as a robotic probe,
- traffic congestion maps (depicting either historical or real time data) and suggested alternative directions,
- information on nearby amenities such as restaurants, fueling stations, and tourist attractions.

GPS may be able to answer:

- the roads or paths available,
- traffic congestion and alternative routes,
- roads or paths that might be taken to get to the destination,
- if some roads are busy (now or historically) the best route to take,
- the location of food, banks, hotels, fuel, airports or other places of interests,
- the shortest route between the two locations,
- the different options to drive on highway or back roads.

Dedicated GPS Navigation Devices

Dedicated devices have various degrees of mobility. *Hand-held*, *outdoor*, or *sport* receivers have replaceable batteries that can run them for several hours, making them suitable for hiking, bicycle touring and other activities far from an electric power source. Their screens are small, and some do not show colour, in part to save power. Some use transflective liquid-crystal displays, allowing use in bright sunlight. Cases are rugged and some are water resistant.

Other receivers, often called *mobile* are intended primarily for use in a car, but have a small rechargeable internal battery that can power them for an hour or two away from the car. Special purpose devices for use in a car may be permanently installed and depend entirely on the automotive electrical system.

The pre-installed embedded software of early receivers did not display maps; 21st century ones commonly show interactive street maps (of certain regions) that may also show points of interest, route information and step-by-step routing directions, often in spoken form with a feature called "text to speech".

Due in part to regulations encouraging mobile phone tracking, including E911, the majority of GPS receivers are built into mobile telephones, with varying degrees of coverage and user accessibility. Commercial navigation software is available for most 21st-century smartphones as well as some Java-enabled phones that allows them to use an internal or external GPS receiver (in the latter case, connecting via serial or Bluetooth). Some phones using assisted GPS (A-GPS) function poorly when out of range of their carrier's cell towers. Others can navigate worldwide with satellite GPS signals as well as a dedicated portable GPS receiver does, upgrading their operation to A-GPS mode when in range. Still others have a hybrid positioning system that can use other signals when GPS signals are inadequate.

More bespoke solutions also exist for smartphones with inbuilt GPS capabilities. Some such phones can use tethering to double as a wireless modem for a laptop, while allowing GPS-navigation/localisation as well. One such example is marketed by Verizon Wireless in the United States, and is called VZ Navigator. The system uses gpsOne technology to determine the location, and then uses the mobile phone's data connection to download maps and calculate navigational routes. Other products including iPhone are used to provide similar services. Nokia gives Ovi Maps free on its smartphones and maps can be preloaded.

According to market research from the independent analyst firm Berg Insight, the sales of GPS-enabled GSM/WCDMA handsets was 150 million units in 2009, while only 40 million separate GPS receivers were sold.

GPS navigation applications for mobile phones include on-line (e.g. Waze, Google Maps Navigation) and off-line (e.g. iGo for Android, Maverick and Here (Nokia) for Windows Phone) navigation applications. Google Maps Navigation, which is included with Android, means most smartphone users only need their phone to have a personal navigation assistant.

Many Android smartphones have an additional GPS feature, called EPO (Extended Prediction Orbit). The phone downloads a file to help it locate GPS satellites more quickly and reduce the Time To First Fix.

Palm, Pocket and Laptop PC

Software companies have made available GPS navigation software programmes for in-vehicle use on laptop computers. Benefits of GPS on a laptop include larger map overview, ability to use the keyboard to control GPS functions, and some GPS software for laptops offers advanced trip-planning features not available on other platforms, such as midway stops, capability of finding alternative scenic routes as well as only highway option.

Palms and Pocket PC's can also be equipped with GPS navigation. A pocket PC differs from a dedicated navigation device as it has an own operating system and can also run other applications.

GPS modules

Other GPS devices need to be connected to a computer in order to work. This computer can be a home computer, laptop, PDA, digital camera, or smartphones. Depending on the type of computer and available connectors, connections can be made through a serial or USB cable, as well as Bluetooth, CompactFlash, SD, PCMCIA and the newer ExpressCard. Some PCMCIA/ExpressCard GPS units also include a wireless modem.

Devices usually do not come with pre-installed GPS navigation software, thus, once purchased, the user must install or write their own software. As the user can choose which software to use, it can be better matched to their personal taste. It is very common for a PC-based GPS receiver to come bundled with a navigation software suite. Also, GPS modules are significantly cheaper than complete stand-alone systems (around €50 to €100). The software may include maps only for a particular region, or the entire world, if software such as Google Maps, Networks in Motion's AtlasBook mobile navigation platform, etc., are used.

Privacy Concern

Due to the popularity of GPS devices, privacy of the user becomes a subject of debate. This is because GPS devices can give geo-location information of the user. This is considered as private information and nobody should violate private information without legal approval. However, there were several incidents where the privacy of GPS devices was questioned.

Advertisement

Since GPS devices can give the user's exact location, this helps advertising agents to give more relevant advertisement to the users based on their current location. The agencies might promote shops which are nearby to the users, rather than totally irrelevant shops. The advertising agency also will store the user's location for the agency's future uses. However, the regulatory agents all around the world (especially USA and Europe) start to consider whether geo-location data should be a sensitive data or not. If the data is sensitive data, the marketing team of an agency can not store geo-location of people since this amounts to a privacy violation. However, if the regulatory agents choose to consider geo-location as non-sensitive data, then private companies can have permission to store the user's location in their database.

Surveillance

Privacy concerns also arise when employers use GPS tracking units to track their employees' location, for example using vehicle tracking systems. This raises a major question about whether this violates personal privacy of employees. It raises a lot more concern for privacy violation if the employers collect geo-location data of their employee after work hours and during their holidays. In 2010, New York Civil Liberties Union filed a case against the Labor Department for firing Michael Cunningham after tracking Michael Cunningham's daily activity and locations using a GPS device that was attached to his car. This raises a few questions regarding the limit of surveillance. The worst privacy violation was committed by the FBI when they tracked down Antoine Jones's GPS devices, even without any search warrants. Later the Federal Appeal Court rejected the FBI's surveillance data as evidence against Antoine Jones.

Stalking

GPS devices are also used by private investigators in order to give more information to their clients. They will plant their own GPS devices in order to know more about their target. Moreover, some rental car services use the same technique to prevent their customers from going out of their targeted area. They charge additional fees for those who violate their rules. They get this information by using the car's GPS devices.

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Higgs Boson: The God Particle in Physics

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The Higgs boson or Higgs particle is an elementary particle initially theorised in 1964, whose discovery was announced at CERN on 4 July 2012. The discovery has been called "monumental" because it appears to confirm the existence of the Higgs field, which is pivotal to the Standard Model and other theories within particle physics. It would explain why some fundamental particles have mass when the symmetries controlling their interactions should require them to be massless, and why the weak force has a much shorter range than the electromagnetic force. The discovery of a Higgs boson should allow physicists to finally validate the last untested area of the Standard Model's approach to fundamental particles and forces, guide other theories and discoveries in particle physics, and potentially lead to developments in "new" physics.

This unanswered question in fundamental physics is of such importance that it led to a search of more than 40 years for the Higgs boson and finally the construction of one of the world's most expensive and complex experimental facilities to date, the Large Hadron Collider, able to create Higgs bosons and other particles for observation and study. On 4 July 2012, it was announced that a previously unknown particle with a mass between 125 and 127 GeV/c^2 (134.2 and 136.3 amu) had been detected; physicists suspected at the time that it was the Higgs boson. By March 2013, the particle had been proven to behave, interact and decay in many of the ways predicted by the Standard Model, and was also tentatively confirmed to have positive parity and zero spin, two fundamental attributes of a Higgs boson. This appears to be the first elementary scalar particle discovered in nature. More data are needed to determine whether the particle discovered exactly matches the predictions of the Standard Model, or whether, as predicted by some theories, multiple Higgs bosons exist.

The Higgs boson is named after Peter Higgs, one of six physicists who, in 1964, proposed the mechanism that suggested the existence of such a particle. Although Higgs's name has come to be associated with this theory, several researchers between about 1960 and 1972 each independently developed different parts of it. In mainstream media the Higgs boson has often been called the "God particle", from a 1993 book on the topic; the nickname is strongly disliked by many physicists, including Higgs, who regard it as inappropriate sensationalism. In 2013 two of the original researchers, Peter Higgs and François Englert, were awarded the Nobel Prize in Physics for their work and prediction. Englert's co-researcher Robert Brout had died in 2011 and the Nobel is not given posthumously except in unusual circumstances.

In the Standard Model, the Higgs particle is a boson with no spin, electric charge, or color charge. It is also very unstable, decaying into other particles almost immediately. It is a quantum excitation of one of the four components of the Higgs field. The latter constitutes a scalar field, with two neutral and two electrically charged components, and forms a complex doublet of the weak isospin SU(2) symmetry. The field has a "Mexican hat" shaped potential with nonzero strength everywhere (including otherwise empty space), which in its vacuum state breaks the weak isospin symmetry of the electroweak interaction. When this happens, three components of the Higgs field are "absorbed" by the SU(2) and U(1) gauge bosons (the "Higgs mechanism") to become the longitudinal components of the now-massive W and Z bosons of the weak force. The remaining electrically neutral component separately couples to other particles known as fermions (via Yukawa couplings), causing these to acquire mass as well. Some versions of the theory predict more than one kind of Higgs fields and bosons. Alternative "Higgsless" models would have been considered if the Higgs boson were not discovered.

In particle physics, elementary particles and forces give rise to the world around us. Nowadays, physicists explain the behaviour of these particles and how they interact using the Standard Model—a widely accepted and "remarkably" accurate framework based on gauge invariance and symmetries, believed to explain almost everything in the world we see, other than gravity.

But by around 1960 all attempts to create a gauge invariant theory for two of the four fundamental forces had consistently failed at one crucial point: although gauge invariance seemed extremely important, it seemed to make any theory of electromagnetism and the weak force go haywire, by demanding that either many particles with mass were massless or that non-existent forces and massless particles had to exist. Scientists had no idea how to get past this point.

Work done on superconductivity and "broken" symmetries around 1960 led physicist Philip Anderson to suggest in 1962 a new kind of solution that might hold the key. In 1964 a theory was created by 3 different groups of researchers, that showed the problems could be resolved if an unusual kind of field existed throughout the universe. It would cause existing particles to acquire mass instead of new massless particles being formed. By 1972 it had been developed into a comprehensive theory and proved capable of giving "sensible" results. Although there was not yet any proof of such a field, calculations consistently gave answers and predictions that were confirmed by experiments, including very accurate predictions of several other particles, so scientists began to believe this might be true and to search for proof whether or not a Higgs field exists in nature.

If this field did exist, this would be a monumental discovery for science and human knowledge, and is expected to open doorways to new knowledge in many fields. If not, then other more complicated theories would need to be explored. The easiest proof whether or not the field existed was by searching for a new kind of particle it would have to give off, known as "Higgs bosons" or the "Higgs particle". These would be extremely difficult to find, so it was only many years later that experimental technology became sophisticated enough to answer the question.

While several symmetries in nature are spontaneously broken through a form of the Higgs mechanism, in the context of the Standard Model the term "Higgs mechanism" almost always means symmetry breaking of the electroweak field. It is considered proven, but the exact cause has been exceedingly difficult to prove. After 50 years, the Higgs boson's existence – apparently proven in 2013 – would finally confirm that the Standard Model is essentially correct and allow further development, while its non-existence would mean that other theories are needed instead.

Various analogies have also been invented to describe the Higgs field and boson, including analogies with well-known symmetry breaking effects such as the rainbow and prism, electric fields, ripples, and resistance of macro objects moving through media, like people moving through crowds or some objects moving through syrup or molasses. However, analogies based on simple resistance to motion are inaccurate as the Higgs field does not work by resisting motion.

Scientific Impact

Evidence of the Higgs field and its properties would be extremely significant scientifically, for many reasons. The Higgs boson's importance is largely that it is able to be examined using existing knowledge and experimental technology, as a way to confirm and study the entire Higgs field theory.

As yet, there are no known immediate technological benefits of finding the Higgs particle. However, observers in both media and science point out that when fundamental discoveries are made about our world, their practical uses can take decades to emerge, but are often world-changing when they do. A common pattern for fundamental discoveries is for practical applications to follow later, once the discovery has been explored further, at which point they become the basis for social change and new technologies.

For example, in the first half of the 20th century it was not expected that quantum mechanics would make possible transistors and microchips, mobile phones and computers, lasers and M.R.I. scanners. Radio waves were described by their co-discoverer in 1888 as "an interesting laboratory experiment" with "no useful purpose" whatsoever, and are now used in innumerable ways (radar, weather prediction, medicine, television, wireless computing and emergency response), positrons are used in hospital tomography scans, and special and general relativity, which explain black holes also enable satellite-based GPS and satellite navigation ("satnav"). Electric power generation and transmission, motors, and lighting all stemmed from previous theoretical work on electricity and magnetism; air conditioning and refrigeration resulted from thermodynamics. It is impossible to predict how seemingly esoteric knowledge may affect society in the future.

Other observers highlight technological spin-offs from this and related particle physics activities, which have already brought major developments to society. For example, the World Wide Web as used today was created by physicists working in global collaborations on particle experiments at CERN to share their results, and the results of massive amounts of data produced by the Large Hadron Collider have already led to significant advances in distributed and cloud computing, now well established within mainstream services.

Experimental Search

To produce Higgs bosons, two beams of particles are accelerated to very high energies and allowed to collide within a particle detector. Occasionally, although rarely, a Higgs boson will be created fleetingly as part of the collision byproducts. Because the Higgs boson decays very quickly, particle detectors cannot detect it directly. Instead the detectors register all the decay products (the *decay signature*) and from the data the decay process is reconstructed. If the observed decay products match a possible decay process (known as a *decay channel*) of a Higgs boson, this indicates that a Higgs boson may have been created. In practice, many processes may produce similar decay signatures. Fortunately, the Standard Model precisely predicts the likelihood of each of these, and each known process, occurring. So, if the detector detects more decay signatures consistently matching a Higgs boson than would otherwise be expected if Higgs bosons did not exist, then this would be strong evidence that the Higgs boson exists.

Because Higgs boson production in a particle collision is likely to be very rare (1 in 10 billion at the LHC), and many other possible collision events can have similar decay signatures, the data of hundreds of trillions of collisions needs to be analysed and must "show the same picture" before a conclusion about the existence of the Higgs boson can be reached. To conclude that a new particle has been found, particle physicists require that the statistical analysis of two independent particle detectors each indicate that there is lesser than a one-in-a-million chance that the observed decay signatures are due to just background random Standard Model events—i.e., that the observed number of events is more than 5 standard deviations (sigma) different from that expected if there was no new particle. More collision data allows better confirmation of the physical properties of any new particle observed, and allows physicists to decide whether it is indeed a Higgs boson as described by the Standard Model or some other hypothetical new particle.

To find the Higgs boson, a powerful particle accelerator was needed, because Higgs bosons might not be seen in lower-energy experiments. The collider needed to have a high luminosity in order to ensure enough collisions were seen for conclusions to be drawn. Finally, advanced computing facilities were needed to process the vast amount of data (25 petabytes per year as at 2012) produced by the collisions. For the announcement of 4 July 2012, a new collider known as the Large Hadron Collider was constructed at CERN with a planned eventual collision energy of 14 TeV—over seven times any previous collider—and over 300 trillion (3×10^{14}) LHC proton–proton collisions were analysed by the LHC Computing Grid, the world's largest computing grid, comprising over 170 computing facilities in a worldwide network across 36 countries.

In January 2013, CERN director-general Rolf-Dieter Heuer stated that based on data analysis to date, an answer could be possible 'towards' mid-2013, and the deputy chair of physics at Brookhaven National Laboratory stated in February 2013 that a "definitive" answer might require "another few years" after the collider's 2015 restart. In early March 2013, CERN Research Director Sergio Bertolucci stated that confirming spin-0 was the major remaining requirement to determine whether the particle is at least some kind of Higgs boson.

Confirmation of new particle as a Higgs boson, and current status

On 14 March 2013 CERN confirmed that:

> *"CMS and ATLAS have compared a number of options for the spin-parity of this particle, and these all prefer no spin and positive parity [two fundamental criteria of a Higgs boson consistent with the Standard Model]. This, coupled with the measured interactions of the new particle with other particles, strongly indicates that it is a Higgs boson."*

This also makes the particle the first elementary scalar particle to be discovered in nature.

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Chemical Industry

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The chemical industry comprises the companies that produce industrial chemicals. Central to the modern world economy, it converts raw materials (oil, natural gas, air, water, metals, and minerals) into more than 70,000 different products.

One of the first chemicals to be produced in large amounts through industrial process, was sulphuric acid. In 1736, the pharmacist Joshua Ward developed a process for its production that involved heating saltpeter and allowing the sulfur to oxidize and combine with water. It was the first practical production of sulfuric acid on a large scale. John Roebuck and Samuel Garbett were the first to establish a large scale factory in Prestonpans in 1749, which used leaden condensing chambers for the manufacture of sulphuric acid.

Expansion and Maturation

The late 19th century saw an explosion in both the quantity of production and the variety of chemicals that were manufactured. Large chemical industries also took shape in Germany and later in the United States.

Production of artificial manufactured fertilizer for agriculture was pioneered by Sir John Lawes at his purpose built Rothamsted Research facility. In the 1840s he established large works near London for the manufacture of superphosphate of lime. Processes for the vulcanization of rubber were patented by Charles Goodyear in the US and Thomas Hancock in England in the 1840s. The first synthetic dye was discovered by William Henry Perkin in London. He partly transformed aniline into a crude mixture which, when extracted with alcohol, produced a substance with an intense purple colour. He also developed the first synthetic perfumes. However, it was German industry that quickly began to dominate the field of synthetic dyes. The three major firms BASF, Bayer and Hoechst produced several hundred different dyes, and by 1913, the German industry produced almost 90 percent of the world supply of dyestuffs and sold about 80 percent of their production abroad.

The petrochemical industry can be traced back to the oil works of James Young in Scotland and Abraham Pineo Gesner in Canada. The first plastic was invented by Alexander Parkes, an English metallurgist. In 1856, he patented Parkesine, a celluloid based on nitrocellulose treated with a variety of solvents. This material, exhibited at the 1862 London International Exhibition, anticipated many of the modern aesthetic and utility uses of plastics. The industrial production of soap from vegetable oils was started by William Lever and his brother James in 1885 in Lancashire based on a modern chemical process invented by William Hough Watson that used glycerin and vegetable oils.

By the 1920s, chemical firms consolidated into large conglomerates; IG Farben in Germany, Rhône-Poulenc in France and Imperial Chemical Industries in Britain. Dupont became a major chemicals firm in the early 20th century in America.

Products

Polymers and plastics, especially polyethylene, polypropylene, polyvinyl chloride, polyethylene terephthalate, polystyrene and polycarbonate comprise about 80% of the industry's output worldwide. Chemicals are used to make a wide variety of consumer goods, as well as thousands of inputs to agriculture, manufacturing, construction, and service industries. The chemical industry itself consumes 26 percent of its own output. Major industrial customers include rubber and plastic products, textiles, apparel, petroleum refining, pulp and paper, and primary metals. Chemicals are nearly a $3 trillion global enterprise, and the EU and U.S. chemical companies are the world's largest producers.[1]

Sales of the chemical business can be divided into a few broad categories, including basic chemicals (about 35 to 37 percent of the dollar output), life sciences (30 percent), specialty chemicals (20 to 25 percent) and consumer products (about 10 percent).

Basic Chemicals and Commodity Chemicals to Polymers and Speciality Chemicals

Basic chemicals or "commodity chemicals" are a broad chemical category including polymers, bulk petrochemicals and intermediates, other derivatives and basic industrials, inorganic chemicals, and fertilizers. Typical growth rates for basic chemicals are about 0.5 to 0.7 times GDP. Product prices are generally less than fifty cents per pound.

Polymers, the largest revenue segment at about 33 percent of the basic chemicals dollar value, includes all categories of plastics and man-made fibers. The major markets for plastics are packaging, followed by home construction, containers, appliances, pipe, transportation, toys, and games.

Chemicals in the bulk petrochemicals and intermediates are primarily made from liquefied petroleum gas (LPG), natural gas, and crude oil. Their sales volume is close to 30 percent of overall basic chemicals. Typical large-volume products include ethylene, propylene, benzene, toluene, xylenes, methanol, vinyl chloride monomer (VCM), styrene, butadiene, and ethylene oxide. These basic or commodity chemicals are the starting materials used to manufacture many polymers and other more complex organic chemicals particularly those that are made for use in the specialty chemicals category.

Other derivatives and basic industrials include synthetic rubber, surfactants, dyes and pigments, turpentine, resins, carbon black, explosives, and rubber products and contribute about 20 percent of the basic chemicals' external sales.

Inorganic chemicals (about 12 percent of the revenue output) make up the oldest of the chemical categories. Products include salt, chlorine, caustic soda, soda ash, acids (such as nitric acid, phosphoric acid, and sulfuric acid), titanium dioxide, and hydrogen peroxide.

Fertilizers are the smallest category (about 6 percent) and include phosphates, ammonia, and potash chemicals.

Life Sciences

Life sciences (about 30 percent of the dollar output of the chemistry business) include differentiated chemical and biological substances, pharmaceuticals, diagnostics, animal health products, vitamins, and pesticides. While much smaller in volume than other chemical sectors, their products tend to have very high prices—over ten dollars per pound—growth rates of 1.5 to 6 times GDP, and research and development spending at 15 to 25 percent of sales. Life science products are usually produced with very high specifications and are closely scrutinized by government agencies such as the Food and Drug Administration. Pesticides, also called "crop protection chemicals", are about 10 percent of this category and include herbicides, insecticides, and fungicides.

Specialty Chemicals

Specialty chemicals are a category of relatively high valued, rapidly growing chemicals with diverse end product markets. Typical growth rates are one to three times GDP with prices over a dollar per pound. They are generally characterized by their innovative aspects. Products are sold for what they can do rather than for what chemicals they contain. Products include electronic chemicals, industrial gases, adhesives and sealants as well as coatings, industrial and institutional cleaning chemicals, and catalysts. In 2012, excluding fine chemicals, the $546 billion global speciality chemical market was 33% Paints, Coating and Surface

Treatments, 27% Advanced Polymer, 14% Adhesives and Sealants, 13% additives and 13% pigments and inks.

Speciality chemicals are sold as effect or performance chemicals sometimes they are mixtures of formulations unlike “fine chemicals” which are almost always single molecule products.

Consumer products

Consumer products include direct product sale of chemicals such as soaps, detergents, and cosmetics. Typical growth rates are 0.8 to 1.0 times GDP.

Consumers rarely if ever come into contact with basic chemicals but polymers and speciality chemicals are the materials that they will encounter everywhere in their every-day lives, such as in plastics, cleaning materials, cosmetics, paints & coatings, electronic gadgets, automobiles and the materials used to construct their homes. These speciality products are marketed by chemical companies to the downstream manufacturing industries as pesticides, speciality polymers, electronic chemicals, surfactants, construction chemicals, Industrial Cleaners, flavours and fragrances, speciality coatings, printing inks, water soluble polymers, food additives, paper chemicals, oil field chemicals, plastic adhesives, adhesives and sealants, cosmetic chemicals, water management chemicals, catalysts, textile chemicals. Chemical companies rarely supply these products directly to the consumer.

Every year, the American Chemistry Council tabulates the U.S. production volume of the top 100 basic chemicals. In 2000, the aggregate production volume of the top 100 chemicals totalled 502 million tons, up from 397 million tons in 1990. Inorganic chemicals tend to be the largest volume, though much smaller in dollar revenue terms due to their low prices. The top 11 of the 100 chemicals in 2000 were sulfuric acid (44 million tons), nitrogen (34), ethylene (28), oxygen (27), lime (22), ammonia (17), propylene (16), polyethylene (15), chlorine (13), phosphoric acid (13) and diammonium phosphates (12).

Companies

The largest corporate producers worldwide, each with plants in numerous countries, include BASF, Bayer, Ferro, Solvay, Braskem, Celanese/Ticona, Arkema, Degussa, Dow, DuPont, Eastman Chemical Company, ExxonMobil, Givaudan, INEOS, LG Chem, LyondellBasell, Mitsubishi, Monsanto, PPG Industries, SABIC, LANXESS, Shell, and Wanhua along with thousands of smaller firms.

In the U.S. there are 170 major chemical companies. They operate internationally with more than 2,800 facilities outside the U.S. and 1,700 foreign subsidiaries or affiliates operating. The U.S. chemical output is $750 billion a year. The U.S. industry records large trade surpluses and employs more than a million people in the United States alone. The chemical industry is also the second largest consumer of energy in manufacturing and spends over $5 billion annually on pollution abatement.

In Europe the chemical, plastics and rubber sectors are among the largest industrial sectors. Together they generate about 3.2 million jobs in more than 60,000 companies. Since 2000 the chemical sector alone has represented 2/3 of the entire manufacturing trade surplus of the EU.

In 2012 the chemical sector accounted for 12% of the EU manufacturing industry's added value. Europe remains world's biggest chemical trading region with 43% of the world's exports and 37%of the world's imports, although the latest data shows that Asia is catching up with 34% of the exports and 37% of imports. Even so Europe still has a trading surplus with all regions of the world except Japan and China where in 2011 there was a chemical trade balance. Europe's trade surplus with the rest of the world today amounts to 41.7 billion Euros.

Over the 20 years between 1991 and 2011 the European Chemical industry saw its sales increase 295 billion Euros to 539 billion Euros a picture of constant growth. Despite this the European industry's share of the world chemical market has fallen from 36% to 20%. This has resulted from the huge increase production and sales in the emerging markets like India and China. The data suggest that 95% of this impact is from China alone. In 2012 the data from the European Chemical Industry Council (CEFIC)shows that 5 European countries account for 71% of the EU's chemicals sales. These are Germany, France, United Kingdom, Italy and the Netherlands.

The industry includes manufacturers of inorganic- and organic-industrial chemicals, ceramic products, petrochemicals, agrochemicals, polymers and rubber (elastomers), oleochemicals (oils, fats, and waxes), explosives, fragrances and flavors.

Although the pharmaceutical industry is often considered a chemical industry , it has many different characteristics that puts it in a separate category. Other closely related industries include petroleum, glass, paint, ink, sealant, adhesive, and food processing manufacturers.

Though the business of chemistry is worldwide in scope, the bulk of the world's $3.7 trillion chemical output is accounted for by only a handful of industrialized nations. The United States alone produced $689 billion, 18.6 percent of the total world chemical output in 2008.

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Regression Analysis

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In statistics, regression analysis is a statistical technique for estimating the relationships among variables. It includes many techniques for modelling and analyzing several variables, when the focus is on the relationship between a dependent variable and one or more independent variables. More specifically, regression analysis helps one understand how the typical value of the dependent variable changes when any one of the independent variables is varied, while the other independent variables are held fixed. Most commonly, regression analysis estimates the conditional expectation of the dependent variable given the independent variables – that is, the average value of the dependent variable when the independent variables are fixed. Less commonly, the focus is on a quantile, or other location parameter of the conditional distribution of the dependent variable given the independent variables. In all cases, the estimation target is a function of the independent variables called the regression function. In regression analysis, it is also of interest to characterize the variation of the dependent variable around the regression function, which can be described by a probability distribution.

Regression analysis is widely used for prediction and forecasting, where its use has substantial overlap with the field of machine learning. Regression analysis is also used to understand which among the independent variables are related to the dependent variable, and to explore the forms of these relationships. In restricted circumstances, regression analysis can be used to infer causal relationships between the independent and dependent variables. However this can lead to illusions or false relationships, so caution is advisable; for example, correlation does not imply causation.

A large body of techniques for carrying out regression analysis has been developed. Familiar methods such as linear regression and ordinary least squares regression are parametric, in that the regression function is defined in terms of a finite number of unknown parameters that are estimated from the data. Nonparametric regression refers to techniques that allow the regression function to lie in a specified set of functions, which may be infinite-dimensional.

The performance of regression analysis methods in practice depends on the form of the data generating process, and how it relates to the regression approach being used. Since the true form of the data-generating process is generally not known, regression analysis often depends to some extent on making assumptions about this process. These assumptions are sometimes testable if many data are available. Regression models for prediction are often useful even when the assumptions are moderately violated, although they may not perform optimally. However, in many applications, especially with small effects or questions of causality based on observational data, regression methods can give misleading results.

History

The earliest form of regression was the method of least squares, which was published by Legendre in 1805, and by Gauss in 1809. Legendre and Gauss both applied the method to the problem of determining, from astronomical observations, the orbits of bodies about the Sun (mostly comets, but also later the then newly discovered minor planets). Gauss published a further development of the theory of least squares in 1821, including a version of the Gauss–Markov theorem.

The term "regression" was coined by Francis Galton in the nineteenth century to describe a biological phenomenon. The phenomenon was that the heights of descendants of tall ancestors tend to regress down towards a normal average (a phenomenon also known as regression toward the mean). For Galton, regression had only this biological meaning, but his work was later extended by Udny Yule and Karl Pearson to a more general statistical context. In the work of Yule and Pearson, the joint distribution of the response and explanatory variables is assumed to be Gaussian. This assumption was weakened by R.A. Fisher in his works of 1922 and 1925. Fisher assumed that the conditional distribution of the response variable is Gaussian, but the joint distribution need not be. In this respect, Fisher's assumption is closer to Gauss's formulation of 1821.

In the 1950s and 1960s, economists used electromechanical desk calculators to calculate regressions. Before 1970, it sometimes took up to 24 hours to receive the result from one regression.

Regression methods continue to be an area of active research. In recent decades, new methods have been developed for robust regression, regression involving correlated responses such as time series and growth curves, regression in which the predictor or response variables are curves, images, graphs, or other complex data objects, regression methods accommodating various types of missing data, nonparametric regression, Bayesian methods for regression, regression in which the predictor variables are measured with error, regression with more predictor variables than observations, and causal inference with regression.

Regression Models

Regression models involve the following variables:

- The unknown parameters, denoted as β, which may represent a scalar or a vector.
- The independent variables, X.
- The dependent variable, Y.

In various fields of application, different terminologies are used in place of dependent and independent variables.

A regression model relates Y to a function of X and β.

$$Y \approx f(X, \beta)$$

The approximation is usually formalized as $E(Y \mid X) = f(X, \beta)$. To carry out regression analysis, the form of the function f must be specified. Sometimes the form of this function is based on knowledge about the relationship between Y and X that does not rely on the data. If no such knowledge is available, a flexible or convenient form for f is chosen.

Assume now that the vector of unknown parameters β is of length k. In order to perform a regression analysis the user must provide information about the dependent variable Y:

- If N data points of the form (Y,X) are observed, where $N < k$, most classical approaches to regression analysis cannot be performed: since the system of equations defining the regression model is underdetermined, there are not enough data to recover β.
- If exactly $N = k$ data points are observed, and the function f is linear, the equations $Y = f(X, \beta)$ can be solved exactly rather than approximately. This reduces to solving a set of N equations with N unknowns (the elements of β), which has a unique solution as long as the X are linearly independent. If f is nonlinear, a solution may not exist, or many solutions may exist.
- The most common situation is where $N > k$ data points are observed. In this case, there is enough information in the data to estimate a unique value for β that best fits the data in some sense, and the regression model when applied to the data can be viewed as an overdetermined system in β.

In the last case, the regression analysis provides the tools for:

1. Finding a solution for unknown parameters β that will, for example, minimize the distance between the measured and predicted values of the dependent variable Y (also known as method of least squares).
2. Under certain statistical assumptions, the regression analysis uses the surplus of information to provide statistical information about

the unknown parameters β and predicted values of the dependent variable *Y*.

Necessary Number of Independent Measurements

Consider a regression model which has three unknown parameters, β_0, β_1, and β_2. Suppose an experimenter performs 10 measurements all at exactly the same value of independent variable vector X (which contains the independent variables X_1, X_2, and X_3). In this case, regression analysis fails to give a unique set of estimated values for the three unknown parameters; the experimenter did not provide enough information. The best one can do is to estimate the average value and the standard deviation of the dependent variable *Y*. Similarly, measuring at two different values of X would give enough data for a regression with two unknowns, but not for three or more unknowns. If the experimenter had performed measurements at three different values of the independent variable vector X, then regression analysis would provide a unique set of estimates for the three unknown parameters in β. In the case of general linear regression, the above statement is equivalent to the requirement that the matrix X^TX is invertible.

Statistical Assumptions

When the number of measurements, *N*, is larger than the number of unknown parameters, *k*, and the measurement errors ε_i are normally distributed then *the excess of information* contained in $(N - k)$ measurements is used to make statistical predictions about the unknown parameters. This excess of information is referred to as the degrees of freedom of the regression.

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Essence of Botany

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Bud

In botany, a bud is an undeveloped or embryonic shoot and normally occurs in the axil of a leaf or at the tip of the stem. Once formed, a bud may remain for some time in a dormant condition, or it may form a shoot immediately. Buds may be specialized to develop flowers or short shoots, or may have the potential for general shoot development. The term bud is also used in zoology, where it refers to an outgrowth from the body which can develop into a new individual.

Figure: *Flower buds have not yet bloomed into a full-size flower*

Overview

The buds of many woody plants, especially in temperate or cold climates, are protected by a covering of modified leaves called *scales* which tightly enclose the more delicate parts of the bud. Many bud scales are covered by a gummy substance which serves as added protection. When the bud develops, the scales may enlarge somewhat but usually just drop off, leaving on the surface of the growing stem a series of horizontally-elongated scars.

By means of these scars one can determine the age of any young branch, since each year's growth ends in the formation of a bud, the formation of which produces an additional group of bud scale scars. Continued growth of the branch causes these scars to be obliterated after a few years so that the total age of older branches cannot be determined by this means.

Figure: *On the left is an opening influorescence bud, that will develop like the one to the right.*

In many plants scales are not formed over the bud, which is then called a naked bud. The minute underdeveloped leaves in such buds are often excessively hairy. Naked buds are found in some shrubs, like some species of the Sumac and Viburnums (*Viburnum alnifolium* and *V. lantana*) and in herbaceous plants.

In many of the latter, buds are even more reduced, often consisting of undifferentiated masses of cells in the axils of leaves. A terminal bud occurs on the end of a stem and lateral buds are found on the side. A head of cabbage is an exceptionally large terminal bud, while Brussels sprouts are large lateral buds. Since buds are formed in the axils of leaves, their distribution on the stem is the same as that of leaves. There are alternate, opposite, and whorled buds, as well as the terminal bud at the tip of the stem. In many plants buds appear in unexpected places: these are known as adventitious buds.

Often it is possible to find a bud in a remarkable series of gradations of bud scales. In the buckeye, for example, one may see a complete gradation from the small brown outer scale through larger scales which on unfolding become somewhat green to the inner scales of the bud, which are remarkably leaf-like. Such a series suggests that the scales of the bud are in truth leaves, modified to protect the more delicate parts of the plant during unfavourable periods.

Types of Buds

Buds are often useful in the identification of plants, especially for woody plants in winter when leaves have fallen. Buds may be classified and described according to different criteria: location, status, morphology, and function.

Botanists commonly use the following terms:

- for location:
  - o terminal, when located at the tip of a stem (apical is equivalent but rather reserved for the one at the top of the plant);
  - o axillary, when located in the axil of a leaf (lateral is the equivalent but some adventitious buds may be lateral too);
  - o adventitious, when occurring elsewhere, for example on trunk or on roots (some adventitious buds may be former axillary ones reduced and hidden under the bark, other adventitious buds are completely new formed ones).
- for status:
  - o accessory, for secondary buds formed besides a principal bud (axillary or terminal);
  - o resting, for buds that form at the end of a growth season, which will lie dormant until onset of the next growth season;
  - o dormant or latent, for buds whose growth has been delayed for a rather long time. The term is usable as a synonym of *resting*, but is rather employed for buds waiting undeveloped for years, for example epicormic buds;
  - o pseudoterminal, for an axillary bud taking over the function of a terminal bud (characteristic of species whose growth is sympodial: terminal bud dies and is replaced by the closer axillary bud, for examples beech, persimmon, *Platanus* have sympodial growth).
- for morphology:
  - o naked, when not covered by scales;
  - o scaly or covered, when scales (which are in fact transformed and reduced leaves) cover and protect the embryonic parts;
  - o hairy, when also protected by hairs (it may apply either to scaly or to naked buds).
- for function:
  - o vegetative, if only containing vegetative pieces: embryonic shoot with leaves (a leaf bud is the same);
  - o reproductive, if containing embryonic flower(s) (a flower bud is the same);
  - o mixed, if containing both embryonic leaves and flowers.

Within Zoology

The term bud (as in budding) is used by analogy within zoology as well, where it refers to an outgrowth from the body which develops into a new individual. It is a form of asexual reproduction limited to animals or plants of relatively simple structure. In this process a portion of the wall of the parent cell softens and pushes out. The protuberance thus formed enlarges rapidly while at this time the nucleus of the parent cell divides. One of the resulting nuclei passes into the bud, and then the bud is cut off from its parent cell and the process is repeated. Often the daughter cell will begin to bud before it becomes separated from the parent, so that whole colonies of adhering cells may be formed. Eventually cross walls cut off the bud from the original cell.

Chlorophyll

Chlorophyll (also chlorophyl) is a green pigment found in almost all plants, algae, and cyanobacteria. Chlorophyll is an extremely important biomolecule, critical in photosynthesis, which allows plants to absorb energy from light. Chlorophyll absorbs light most strongly in the blue portion of the electromagnetic spectrum, followed by the red portion. However, it is a poor absorber of green and near-green portions of the spectrum, hence the green colour of chlorophyll-containing tissues. Chlorophyll was first isolated by Joseph Bienaimé Caventou and Pierre Joseph Pelletier in 1817.

Chlorophyll and Photosynthesis

Chlorophyll is vital for photosynthesis, which allows plants to absorb energy from light. Chlorophyll molecules are specifically arranged in and around photosystems that are embedded in the thylakoid membranes of chloroplasts. In these complexes, chlorophyll serves two primary functions. The function of the vast majority of chlorophyll (up to several hundred molecules per photosystem) is to absorb light and transfer that light energy by resonance energy transfer to a specific chlorophyll pair in the reaction centre of the photosystems.

The two currently accepted photosystem units are Photosystem II and Photosystem I, which have their own distinct reaction centre chlorophylls, named P680 and P700, respectively. These pigments are named after the wavelength (in nanometers) of their red-peak absorption maximum. The identity, function and spectral properties of the types of chlorophyll in each photosystem are distinct and determined by each other and the protein structure surrounding them. Once extracted from the protein into a solvent (such as acetone or methanol), these chlorophyll pigments can be separated in a simple paper chromatography experiment and, based on the number of polar groups between chlorophyll a and chlorophyll b, will chemically separate out on the paper.

The function of the reaction centre chlorophyll is to use the energy absorbed by and transferred to it from the other chlorophyll pigments in the photosystems to undergo a charge separation, a specific redox reaction in which the chlorophyll donates an electron into a series of molecular intermediates called an electron transport chain.

The charged reaction centre chlorophyll ($P680^+$) is then reduced back to its ground state by accepting an electron. In Photosystem II, the electron that reduces $P680^+$ ultimately comes from the oxidation of water into O_2 and H^+ through several intermediates.

This reaction is how photosynthetic organisms such as plants produce O_2 gas, and is the source for practically all the O_2 in Earth's atmosphere. Photosystem I typically works in series with Photosystem II; thus the $P700^+$ of Photosystem I is usually reduced, via many intermediates in the thylakoid membrane, by electrons ultimately from Photosystem II. Electron transfer reactions in the thylakoid membranes are complex, however, and the source of electrons used to reduce $P700^+$ can vary.

The electron flow produced by the reaction centre chlorophyll pigments is used to shuttle H^+ ions across the thylakoid membrane, setting up a chemiosmotic potential used mainly to produce ATP chemical energy; and those electrons ultimately reduce $NADP^+$ to NADPH, a universal reductant used to reduce CO_2 into sugars as well as for other biosynthetic reductions.

Reaction centre chlorophyll–protein complexes are capable of directly absorbing light and performing charge separation events without other chlorophyll pigments, but the absorption cross section (the likelihood of absorbing a photon under a given light intensity) is small.

Thus, the remaining chlorophylls in the photosystem and antenna pigment protein complexes associated with the photosystems all cooperatively absorb and funnel light energy to the reaction centre. Besides chlorophyll *a*, there are other pigments, called accessory pigments, which occur in these pigment–protein antenna complexes. A green sea slug, *Elysia chlorotica*, has been found to use the chlorophyll it has eaten to perform photosynthesis for itself. This process is known as kleptoplasty, and no other animal has been found to have this ability.

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Environmental Chemistry

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Air Pollution

Air pollution is the introduction into the atmosphere of chemicals, particulates, or biological materials that cause discomfort, disease, or death to humans, damage other living organisms such as food crops, or damage the natural environment or built environment.

The atmosphere is a complex dynamic natural gaseous system that is essential to support life on planet Earth. Stratospheric ozone depletion due to air pollution has long been recognized as a threat to human health as well as to the Earth's ecosystems.

Indoor air pollution and urban air quality are listed as two of the World's Worst Toxic Pollution Problems in the 2008 Blacksmith Institute World's Worst Polluted Places report.

Pollutants

A substance in the air that can be adverse to humans and the environment is known as an air pollutant. Pollutants can be in the form of solid particles, liquid droplets, or gases. In addition, they may be natural or man-made. Pollutants can be classified as primary or secondary. Usually, primary pollutants are directly produced from a process, such as ash from a volcanic eruption, the carbon monoxide gas from a motor vehicle exhaust or sulphur dioxide released from factories. Secondary pollutants are not emitted directly. Rather, they form in the air when primary pollutants react or interact. An important example of a secondary pollutant is ground level ozone — one of the many secondary pollutants that make up photochemical smog. Some pollutants may be both primary and secondary: that is, they are both emitted directly and formed from other primary pollutants.

Major primary pollutants produced by human activity include:

- Sulphur oxides (SO_x) - especially sulphur dioxide, a chemical compound with the formula SO_2. SO_2 is produced by volcanoes and in various industrial processes. Since coal and petroleum often contain sulphur compounds, their combustion generates sulphur dioxide. Further oxidation of SO_2, usually in the presence of a catalyst such as NO_2, forms H_2SO_4, and thus acid rain.[2] This is one of the causes for concern over the environmental impact of the use of these fuels as power sources.

- Nitrogen oxides (NO_x) - especially nitrogen dioxide are expelled from high temperature combustion, and are also produced naturally during thunderstorms by electric discharge. Can be seen as the brown haze dome above or plume downwind of cities. Nitrogen dioxide is the chemical compound with the formula NO_2. It is one of the several nitrogen oxides. This reddish-brown toxic gas has a characteristic sharp, biting odor. NO_2 is one of the most prominent air pollutants.
- Carbon monoxide (CO)- is a colourless, odorless, non-irritating but very poisonous gas. It is a product by incomplete combustion of fuel such as natural gas, coal or wood. Vehicular exhaust is a major source of carbon monoxide.
- Volatile organic compounds - VOCs are an important outdoor air pollutant. In this field they are often divided into the separate categories of methane (CH_4) and non-methane (NMVOCs). Methane is an extremely efficient greenhouse gas which contributes to enhanced global warming. Other hydrocarbon VOCs are also significant greenhouse gases via their role in creating ozone and in prolonging the life of methane in the atmosphere, although the effect varies depending on local air quality. Within the NMVOCs, the aromatic compounds benzene, toluene and xylene are suspected carcinogens and may lead to leukemia through prolonged exposure. 1,3-butadiene is another dangerous compound which is often associated with industrial uses.
- Particulates, alternatively referred to as particulate matter (PM), atmospheric particulate matter, or fine particles, are tiny particles of solid or liquid suspended in a gas. In contrast, aerosol refers to particles and the gas together. Sources of particulates can be man made or natural. Some particulates occur naturally, originating from volcanoes, dust storms, forest and grassland fires, living vegetation, and sea spray. Human activities, such as the burning of fossil fuels in vehicles, power plants and various industrial processes also generate significant amounts of aerosols. Averaged over the globe, anthropogenic aerosols—those made by human activities—currently account for about 10 percent of the total amount of aerosols in our atmosphere. Increased levels of fine particles in the air are linked to health hazards such as heart disease, altered lung function and lung cancer.

- Persistent free radicals connected to airborne fine particles could cause cardiopulmonary disease.
- Toxic metals, such as lead and mercury, especially their compounds.
- Chlorofluorocarbons (CFCs) - harmful to the ozone layer emitted from products currently banned from use.
- Ammonia (NH_3) - emitted from agricultural processes. Ammonia is a compound with the formula NH_3. It is normally encountered as a gas with a characteristic pungent odor. Ammonia contributes significantly to the nutritional needs of terrestrial organisms by serving as a precursor to foodstuffs and fertilizers. Ammonia, either directly or indirectly, is also a building block for the synthesis of many pharmaceuticals. Although in wide use, ammonia is both caustic and hazardous.
- Odors — such as from garbage, sewage, and industrial processes
- Radioactive pollutants - produced by nuclear explosions, nuclear events, war explosives, and natural processes such as the radioactive decay of radon.

Secondary pollutants include:

- Particulates created from gaseous primary pollutants and compounds in photochemical smog. Smog is a kind of air pollution; the word "smog" is a portmanteau of smoke and fog. Classic smog results from large amounts of coal burning in an area caused by a mixture of smoke and sulphur dioxide. Modern smog does not usually come from coal but from vehicular and industrial emissions that are acted on in the atmosphere by ultraviolet light from the sun to form secondary pollutants that also combine with the primary emissions to form photochemical smog.
- Ground level ozone (O_3) formed from NO_x and VOCs. Ozone (O_3) is a key constituent of the troposphere. It is also an important constituent of certain regions of the stratosphere commonly known as the Ozone layer. Photochemical and chemical reactions involving it drive many of the chemical processes that occur in the atmosphere by day and by night. At abnormally high concentrations brought about by human activities (largely the combustion of fossil fuel), it is a pollutant, and a constituent of smog.
- Peroxyacetyl nitrate (PAN) - similarly formed from NO_x and VOCs.

Minor air pollutants include:

- A large number of minor hazardous air pollutants. Some of these are regulated in USA under the Clean Air Act and in Europe under the Air Framework Directive.

- A variety of persistent organic pollutants, which can attach to particulates.

Persistent organic pollutants (POPs) are organic compounds that are resistant to environmental degradation through chemical, biological, and photolytic processes. Because of this, they have been observed to persist in the environment, to be capable of long-range transport, bioaccumulate in human and animal tissue, biomagnify in food chains, and to have potential significant impacts on human health and the environment.

Sources

Sources of air pollution refer to the various locations, activities or factors which are responsible for the releasing of pollutants into the atmosphere. These sources can be classified into two major categories which are:

Anthropogenic sources (man-made sources) mostly related to burning different kinds of fuel

- "Stationary Sources" include smoke stacks of power plants, manufacturing facilities (factories) and waste incinerators, as well as furnaces and other types of fuel-burning heating devices. In developing and poor countries, traditional biomass burning is the major source of air pollutants; traditional biomass includes wood, crop waste and dung.
- "Mobile Sources" include motor vehicles, marine vessels, aircraft and the effect of sound etc.
- Chemicals, dust and controlled burn practices in agriculture and forestry management. Controlled or prescribed burning is a technique sometimes used in forest management, farming, prairie restoration or greenhouse gas abatement. Fire is a natural part of both forest and grassland ecology and controlled fire can be a tool for foresters. Controlled burning stimulates the germination of some desirable forest trees, thus renewing the forest.
- Fumes from paint, hair spray, varnish, aerosol sprays and other solvents
- Waste deposition in landfills, which generate methane. Methane is highly flammable and may form explosive mixtures with air. Methane is also an asphyxiant and may displace oxygen in an enclosed space. Asphyxia or suffocation may result if the oxygen concentration is reduced to below 19.5% by displacement.
- Military, such as nuclear weapons, toxic gases, germ warfare and rocketry

Natural Sources

- Dust from natural sources, usually large areas of land with little or no vegetation
- Methane, emitted by the digestion of food by animals, for example cattle
- Radon gas from radioactive decay within the Earth's crust. Radon is a colourless, odorless, naturally occurring, radioactive noble gas that is formed from the decay of radium. It is considered to be a health hazard. Radon gas from natural sources can accumulate in buildings, especially in confined areas such as the basement and it is the second most frequent cause of lung cancer, after cigarette smoking.
- Smoke and carbon monoxide from wildfires
- Vegetation, in some regions, emits environmentally significant amounts of VOCs on warmer days. These VOCs react with primary anthropogenic pollutants—specifically, NO_x, SO_2, and anthropogenic organic carbon compounds—to produce a seasonal haze of secondary pollutants.
- Volcanic activity, which produce sulphur, chlorine, and ash particulates

Emission Factors

Air pollutant emission factors are representative values that people attempt to relate the quantity of a pollutant released to the ambient air with an activity associated with the release of that pollutant. These factors are usually expressed as the weight of pollutant divided by a unit weight, volume, distance, or duration of the activity emitting the pollutant (e.g., kilograms of particulate emitted per tonne of coal burned). Such factors facilitate estimation of emissions from various sources of air pollution. In most cases, these factors are simply averages of all available data of acceptable quality, and are generally assumed to be representative of long-term averages.

There are 12 compounds in the list of POPs. Dioxins and furans are two of them and are intentionally created by combustion of organics, like open burning of plastics. The POPs are also endocrine disruptor and can mutate the human genes.

The United States Environmental Protection Agency has published a compilation of air pollutant emission factors for a multitude of industrial sources. The United Kingdom, Australia, Canada and many other countries have published similar compilations, as well as the European Environment Agency.

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The Impact of Mineral Toxicity Stress

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Soil pH or soil reaction is an indication of the acidity or alkalinity of soil and is measured in pH units. Soil pH is defined as the negative logarithm of the hydrogen ion concentration.

The pH scale goes from 0 to 14 with pH 7 as the neutral point. As the amount of hydrogen ions in the soil increases the soil pH decreases thus becoming more acidic. From pH 7 to 0 the soil is increasingly more acidic and from pH 7 to 14 the soil is increasingly more alkaline or basic.

Descriptive terms commonly associated with certain ranges in soil pH are:

- extremely acid, < than 4.5; lemon=2.5; vinegar=3.0; stomach acid=2.0; soda=2–4
- very strongly acid, 4.5–5.0; beer=4.5–5.0; tomatoes=4.5
- strongly acid 5.1–5.5; carrots=5.0; asparagus=5.5; boric acid=5.2; cabbage=5.3
- moderately acid, 5.6–6.0; potatoes=5.6
- slightly acid, 6.1–6.5; salmon=6.2; cow's milk=6.5
- neutral, 6.6–7.3; saliva=6.6–7.3; blood=7.3; shrimp=7.0
- slightly alkaline, 7.4–7.8; eggs=7.6–7.8
- moderately alkaline, 7.9–8.4; sea water=8.2; sodium bicarbonate=8.4
- strongly alkaline, 8.5–9.0; borax=9.0
- very strongly alkaline, > than 9.1; milk of magnesia=10.5, ammonia=11.1; lime=12

Measuring Soil pH

Soil pH provides various clues about soil properties and is easily determined. The most accurate method of determining soil pH is by a pH metre. A second method which is simple and easy but less accurate then using a pH metre, consists of using certain indicators or dyes.

Phosphorus is never readily soluble in the soil but is most available in soil with a pH range centred around 6.5. Extremely and strongly acid soils (pH 4.0-5.0) can have high concentrations of soluble aluminum, iron and manganese which may be toxic to the growth of some plants. A pH range of approximately 6 to 7 promotes the most ready availability of plant nutrients.

But some plants, such as azaleas, rhododendrons, blueberries, white potatoes and conifer trees, tolerate strong acid soils and grow well. Also, some plants do well only in slightly acid to moderately alkaline soils. However, a slightly alkaline (pH 7.4-7.8) or higher pH soil can cause a problem with the availability of iron to pin oak and a few other trees in Central New York causing chlorosis of the leaves which will put the tree under stress leading to tree decline and eventual mortality.

The soil pH can also influence plant growth by its effect on activity of beneficial microorganisms Bacteria that decompose soil organic matter are hindered in strong acid soils. This prevents organic matter from breaking down, resulting in an accumulation of organic matter and the tie up of nutrients, particularly nitrogen, that are held in the organic matter.

Changes in Soil pH

Soils tend to become acidic as a result of: (1) rainwater leaching away basic ions (calcium, magnesium, potassium and sodium); (2) carbon dioxide from decomposing organic matter and root respiration dissolving in soil water to form a weak organic acid; (3) formation of strong organic and inorganic acids, such as nitric and sulfuric acid, from decaying organic matter and oxidation of ammonium and sulfur fertilizers. Strongly acid soils are usually the result of the action of these strong organic and inorganic acids.

Lime is usually added to acid soils to increase soil pH. The addition of lime not only replaces hydrogen ions and raises soil pH, thereby eliminating most major problems associated with acid soils but it also provides two nutrients, calcium and magnesium to the soil. Lime also makes phosphorus that is added to the soil more available for plant growth and increases the availability of nitrogen by hastening the decomposition of organic matter. Liming materials are relatively inexpensive, comparatively mild to handle and leave no objectionable residues in the soil. Some common liming materials are: (1) Calcic limestone which is ground limestone; (2) Dolomitic limestone from ground limestone high in magnesium; and (3) Miscellaneous sources such as wood ashes. The amount of lime to apply to correct a soil acidity problem is affected by a number of factors, including soil pH, texture (amount of sand, silt and clay), structure, and amount of organic matter. In addition to soil variables the crops or plants to be grown influence the amount of lime needed.

Causes and Effects of Soil Acidity

Soil acidity is a crop production problem of increasing concern in central and western Oklahoma. Although acid soil conditions are more widespread in eastern Oklahoma, the more natural occurrence there has resulted in farm operators being better able to manage soil acidity in that part of the state. However, in central and western Oklahoma the problem appears to grow with time. This fact sheet explains why soils become acid and the problems acid soils create for crop production. OSU Extension Facts No. 2229 explains how soil acidity and the lime requirement are determined by soil testing. A subsequent fact sheet discusses managing wheatland soils in Oklahoma.

Why Soils are Acid

The four major causes for soils to become acid arelisted below:

1. Rainfall and leaching
2. Acidic parent material
3. Organic matter decay
4. Harvest of high yielding crops.

The above causes of soil acidity are more easily understood when we consider that a soil is acid when there is an abundance of acidic cations (pronounced cat-eyeon), like hydrogen (H^+) and aluminium (Al^{+++}) present compared to the alkaline cations like calcium (Ca^{++}), magnesium (Mg^{++}), potassium (K^+), and sodium (Na^+).

Rainfall and Leaching

Excessive rainfall is an effective agent for removing basic cations over a long time period (thousands of years). In Oklahoma, for example, we can generally conclude that soils are naturally acidic if the rainfall is above 30 inches per year. Therefore, soils east of I-35 tend to be acidic and those west of I-35, alkaline. There are many exceptions to this rule though, mostly as a result of item 4, intensive crop production. Rainfall is most effective in causing soils to become acidic if a lot of water moves through the soil rapidly. Sandy soils are often the first to become acidic because water percolates rapidly, and sandy soils contain only a small reservoir of bases (buffer capacity) due to low clay and organic matter contents. Since the effect of rainfall on acid soil development is very slow, it may take hundreds of years for new parent material to become acidic under high rainfall.

Parent Material

Due to differences in chemical composition of parent materials, soils will become acidic after different lengths of time. Thus, soils that developed from granite material are likely to be more acidic than soils developed from calcareous shale or limestone.

Organic Matter Decay

Decaying organic matter produces H+ which is responsible for acidity. The carbon dioxide (CO_2) produced by decaying organic matter reacts with water in the soil to form a weak acid called carbonic acid. This is the same acid that develops when CO_2 in the atmosphere reacts with rain to form acid rain naturally.

Several organic acids are also produced by decaying organic matter, but they are also weak acids. Like rainfall, the contribution to acid soil development by decaying organic matter is generally very small, and it would only be the accumulated effects of many years that might ever by measured in a field.

Crop Production

Harvesting of crops has its affect on soil acidity development because crops absorb the lime-like elements, as cations, for their nutrition. When these crops are harvested and the yield is removed from the field, then some of the basic material responsible for counteracting the acidity developed by other processes is lost, and the net affect is increased soil acidity. Increasing crop yields will cause greater amounts of basic material to be removed. Grain contains less basic materials than leaves or stems. For this reason, soil acidity will develop faster under continuous wheat pasture than when grain only is harvested. High yielding forages, such as bermudagrass or alfalfa, can cause soil acidity to develop faster than with other crops.

The approximate amount of lime-like elements removed from the soil by a 30 bushel wheat crop. Note that there is almost four times as much lime material removed in the forage as the grain. This explains why wheat pasture that is grazed out will become acidic much faster than when grain alone is produced. Using 50 percent ECCE lime, it would take about one ton every 10 years to maintain soil pH when straw (or forage) and grain are produced annually at the 30 bushel per acre level.

The use of fertilizers, especially those supplying nitrogen, has often been blamed as a cause of soil acidity. Although acidity is produced when ammonium containing materials are transformed to nitrate in the soil, this is countered by other reactions and the final crop removal of nitrogen in a form similar to that in the fertilizer.

What Happens in Acid Soils

Knowing the soil pH helps identify the kinds of chemical reactions that are likely to be taking place in the soil. Generally, the most important reactions from the standpoint of crop production are those dealing with solubilities of compounds or materials in soils. In this regard, we are most concerned with the affects of pH on the availability of toxic elements and nutrient elements.

Toxic elements like aluminium and manganese are the major causes for crop failure in acid soils. These elements are a problem in acid soils because they are more soluble at low pH. In other words, more of the solid form of these elements will dissolve in water when the pH is acid. There is always a lot of aluminium present in soils because it is a part of most clay particles.

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Mechanics of Fluids

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What is Fluid Mechanics?

One can construct a free body diagram of a little fluid particle to visualize both the normal and shear stresses acting on the body: Free body diagram for a fluidparticle at rest. Consider a tiny fluid element (a very small chunk of the fluid) in a case where the fluid is at rest (or moving at constant speed in a straight line). A fluid at rest can have only normal stresses, since a fluid at rest cannot resist ashear stress. In this case, the sum of all the forces must balance the weight of the fluid element.

This condition is known as hydrostatics. Here, pressure is the only normal stress which exists. Pressure always acts positively inward. Obviously, the pressure at the bottom of the fluid element must be slightly larger than that at the top, in order for the total pressure force to balance the weight of the element. Meanwhile, the pressure at the right face must be equal to that on the left face, so that the sum of forces in the horizontal direction is zero. [Note: This diagram is two-dimensional, but an actual fluid element is three-dimensional.

Consider a tiny fluid element (a very small chunk of the fluid) that is moving around in some flow field. Since the fluid is in motion, it can have both normal and shear stresses, as shown by the free body diagram. The vector sum of all forces acting on the fluid element must equal the mass of the element times its acceleration (Newton's second law). Likewise, the net moment about the centre of the body can be obtained by summing the forces due to each shear stress times its moment arm. Fluids at rest cannot resist a shear stress; in other words, when a shear stress is applied to a fluid at rest, the fluid will not remain at rest, but will move because of the shear stress. For a good illustration of this, consider the comparison of a fluid and a solid under application of a shear stress: A fluid can easily be distinguished from a solid by application of a shear stress, since, by definition, a fluid at rest cannot resist a shearstress. If a shear stress is applied to the surface of a solid, the solid will deform a little, and then remain at rest (in its new distorted shape). One can say that the solid (at rest) is able to resist the shear stress. Now consider a fluid (in a container). When a shearstress is applied to the surface of the fluid, the fluid will continuously deform, i.e. it will set up some kind of flow pattern inside the container. In other words, one can say that the fluid (at rest) is unable to resist the shear stress. That is to say, it can not remain at rest under application of a shear stress.

Properties of Fluid

The top plate will experience a friction force to the left, since it is doing work trying to drag the fluid along with it to the right. The fluid at the top of the channel will experience an equal and opposite force (i.e. to the righ Similarly the bottom plate will experience a friction force to the right, since the fluid is trying to puthe plate along with it to the right. The fluid at the bottom of the channel will feel an equal and opposite force, i.e. to the left. In fluid mechanics, shear stress, defined as a tangential force per unarea, is used rather than force itself, and is commonly denot). Llitted by (Greek letter "tau"). In simple shear flow such as this, the shear stress is directly proportional to the rate of deformation of the fluid, which in this case is equal to the slope of the velocity profile /

Surface Tension and Capillarity

Surface tension is a property of liquids which is felt at the interface between the liquid and another fluid (typically a gas). Surfacetension has dimensions of force per unit length, and always acts parallel to the interface. Surface molecules are subject to an attractive force from nearby surface molecules so that the surface is in a state of tension. A soap bubble is a good example to illustrate the effects of surface tension. How does a soap bubble remain spherical in shape? The answer is that there is a higher pressure inside the bubble than outside, much like a balloon. In fact, surfacetension in the soap film acts much the same as the tension in the skin of a balloon. Consider a soap bubble of radiu can be found by considering the free-body diagram of half a bubble. Note that surface tension acts along the circumference (resulting from cutting across the two interfaces) and the pressure acts on the area of the half-bubble. By statics (to be explained later), the net force due to the pressure is equal to the pressure times the projected area. Hence, balancing the forces due to surface tension and pressure difference:

Surface tension: Surface tension is also important at the interface between a liquid, a gas, and a solid. For example, a meniscus occurs when the surface of a liquid touches a solid wall, as most readily noticed when a capillary tube is placed in a liquid. Consider a glass capillary tube inserted into a liquid, such as water. The water will rise up the tube to a heigh because surface tension pulls the surface of the water towards the glass, as shown. The meniscus is the curved surface at the top of the water column. The height of the water column can be found by summing all forces acting on the water column as a free body diagram. (This is a statics problem since there is no acceleration.) The downward force is due to gravity, i.e. the weight of the water column. The only upward force available to balance the weight is that caused by surface tension (pressure forces all cancel out, as will be explained in a later lecture). Column height the liquid is repelled by the solid, and tries not to ”wet” it. For example, water wets glass, but not wax. Mercury on the other hand does not wet glass.

Vapour Pressure

Vapour pressure is defined as the pressure at which a liquid will boil (vaporize). Vapour pressure rises as temperature rises. For example, suppose you are camping on a high mountain (10,000 ft. orroughly 3,000 m in altitude).

Fluid Mechanics

The atmospheric pressure at this elevation is about 70 kPa. The vapour pressure of water is also around 70 kPa. From this it can be stated that at 10,000 ft. of elevation, water boils at around 90 C at standard sea level pressure.

This has consequences for cooking. For example, eggs have to be cooked longer at elevation to become hard-boiled since they cook at a lower temperature. A pressure cooker has the opposite effect. Namely, the tight lid on a pressure cooker causes the pressure to increase above the normal atmospheric value. This causes water to boil at a temperature even greater than 100. Vapour pressure is important to fluid flows because, in general, pressure in a flow decreases as velocity increases. Cavitation which is generally destructive and undesirable. In particular, at high speeds the local pressure of a liquid sometimes drops below the vapour pressure of the liquid.

In such a case, occurs. In other words, a “cavity” or bubble of vapour appears because the liquid vaporizes or boils at the location where the pressure dips below the local vapour pressure. Cavitation is not desirable for several reasons. First, it causes noise (as the cavitation bubbles collapse when they migrate into regions of higher pressure). Second, it can lead to inefficiencies and reduction of heat transfer in pumps and turbines (turbomachines). Finally, the collapse of these cavitation bubbles causes pitting and corrosion of blades and other surfaces nearby.

That is, air is typically four orders of magnitude more compressible than water. For most practical purposes liquids may be regarded as incompressible. However, there are certain cases, such as unsteady flow in pipes (e.g., water hammer), where the compressibility should be taken into account. Gases may also be treated as incompressible if the change in density is very small (typically less than 3%). An ideal fluid is an incompressible fluid. Pressure disturbances imposed on a fluid move in waves.

These pressure waves move at a velocity equal to that of sound through the fluid. The velocity, or celerity, Compressible flows are inherently complicated because the laws of thermodynamics, as well as the laws of fluid mechanics, operate simultaneously.

Concluding Remarks Fluid mechanics represents that branch of applied mechanics dealing with the behaviour of fluids at rest and in motion In the development of the principles of fluid mechanics, some fluid properties play principal roles, other only minor roles or no roles at all for a particular problem. Influid statics, weight is the important property, whereas in fluid flow, density and viscosity are predominant properties. Where appreciable compressibility occurs, principles of thermodynamics must be considered. Vapour pressure becomes important when low gauge pressures are involved, and surface tension affects static and flow conditions in small passage.

Piezometre Tube

The simplest manometre is a tube, open at the top, which is attached to a vessel or a pipe containing liquid at a pressure (higher than atmospheric) to be measured. This simple device is known as a piezometre tube. As the tube is open to the atmosphere the pressure measured is relative to atmospheric so is gauge pressure:= - Finally, note that in many cases (such as with air pressure being measured by a mercury manometre), the density of manometre fluid 2 is much greater than that of fluid 1.

Differential Manometre

A differential manometre can be used to measure the difference in pressure between two containers or two points in the same system.

Inclined-tube Manometre

As shown above, the differential reading is proportional to the pressure difference. If the pressure difference is very small, the reading may be too small to be measured with good accuracy. To increase the sensitivity of the differential reading, one leg of the manometre can be inclined at an angle.

Applied Mechanics

Applied mechanics is a branch of the physical sciences and the practical application of mechanics. Applied mechanics examines the response of bodies (solids and fluids) or systems of bodies to external forces. Some examples of mechanical systems include the flow of a liquid under pressure, the fracture of a solid from an applied force, or the vibration of an ear in response to sound.

A practitioner of the discipline is known as a mechanician. Applied mechanics, as its name suggests, bridges the gap between physical theory and its application to technology. As such, applied mechanics is used in many fields of engineering, especially mechanical engineering. In this context, it is commonly referred to as engineering mechanics. Much of modern engineering mechanics is based on Isaac Newton's laws of motion while the modern practice of their application can be traced back to Stephen Timoshenko, who is said to be the father of modern engineering mechanics. Within the theoretical sciences, applied mechanics is useful in formulating new ideas and theories, discovering and interpreting phenomena, and developing experimental and computational tools. In the application of the natural sciences, mechanics was said to be complemented by thermodynamics by physical chemists Gilbert N. Lewis and Merle Randall, the study of heat and more generally energy, and electromechanics, the study of electricity and magnetism.

Applied Mechanics in Practice

As a scientific discipline, applied mechanics derives many of its principles and methods from the Physical sciences (in particular, Mechanics and Classical Mechanics), from Mathematics and, increasingly, from Computer Science. As such, Applied Mechanics shares similar methods, theories, and topics with Applied Physics, Applied Mathematics, and Computational Science.

As an enabling discipline, applied mechanics has received impetus from the study of natural phenomena such as orbits of planets, circulation of blood, locomotion of animals, crawling of cells, formation of mountains, and propagation of seismic waves. Such studies have resulted in disciplines such as celestial mechanics, biomechanics and geomechanics.

As a practical discipline, applied mechanics has also advanced by participating in major inventions throughout history, such as buildings, ships, automobiles, railways, petroleum refineries, engines, airplanes, nuclear reactors, composite materials, computers, and medical implants. In such connections, the discipline is also known as Engineering

Mechanics, often practiced within Civil Engineering, Mechanical Engineering, Construction Engineering, Materials Science and Engineering, Aerospace Engineering, Chemical Engineering, Electrical Engineering, Nuclear Engineering, Structural engineering and Bioengineering.

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Modelling of Prime Movers and Generators

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In engineering, a prime mover is an engine that converts fuel to useful work. In locomotives, the prime mover is thus the source of power for its propulsion. The term is generally used when discussing any locomotive powered by an internal combustion engine. The term is also applied to engine-generator sets, where the engine is termed the prime mover, as distinct from the generator.

In a diesel-mechanical locomotive, prime mover refers to the diesel engine that is mechanically coupled to the driving wheels (drivers). In a diesel-electric locomotive, prime mover refers to the diesel engine that rotates the main generator responsible for producing electricity to power the traction motors that are geared to the drivers. The prime mover can also be a gas turbine instead of a diesel engine. In either case, the generator, traction motors and interconnecting apparatus are considered to be the power transmission system and not part of the prime mover. A wired-electric or battery-electric locomotive has no on-board prime mover, instead relying on an external power station.

Power Unit

The term *power unit* is also sometimes used in application to diesel locomotives, with a similar meaning. Where the engine and generator set of a diesel-electric locomotive are removable as a unit, it is usual to term the coupled pair of them as "the power unit", but "prime mover" is applied to the diesel engine alone.

Figure: *Power unit (engine and generator right) from a diesel-electric locomotive*

Electric Generator

In electricity generation, an electric generator is a device that converts mechanical energy to electrical energy. A generator forces electric current to flow through an external circuit. The source of mechanical energy may be a reciprocating or turbine steam engine, water falling through a turbine or waterwheel, an internal combustion engine, a wind turbine, a hand crank, compressed air, or any other source of mechanical energy. Generators provide nearly all of the power for electric power grids.

The reverse conversion of electrical energy into mechanical energy is done by an electric motor, and motors and generators have many similarities. Many motors can be mechanically driven to generate electricity and frequently make acceptable generators. Before the connection between magnetism and electricity was discovered, electrostatic generators were used. They operated on electrostatic principles. Such generators generated very high voltage and low current. They operated by using moving electrically charged belts, plates, and disks that carried charge to a high potential electrode. The charge was generated using either of two mechanisms:

- Electrostatic induction
- The triboelectric effect, where the contact between two insulators leaves them charged.

Because of their inefficiency and the difficulty of insulating machines that produced very high voltages, electrostatic generators had low power ratings, and were never used for generation of commercially significant quantities of electric power. The Wimshurst machine and Van de Graaff generator are examples of these machines that have survived.

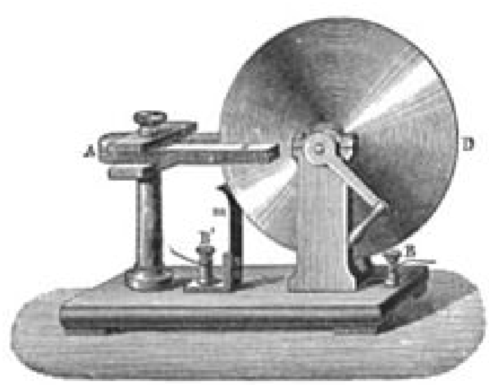

Figure: *Faraday disk, the first electric generator. The horseshoe-shaped magnet* (A) *created a magnetic field through the disk* (D). *When the disk was turned, this induced an electric current radially outward from the centre toward the rim. The current flowed out through the sliding spring contact* m, *through the external circuit, and back into the centre of the disk through the axle.*

In 1827, Hungarian Anyos Jedlik started experimenting with the electromagnetic rotating devices which he called electromagnetic self-rotors, now called the Jedlik's dynamo.

In the prototype of the single-pole electric starter (finished between 1852 and 1854) both the stationary and the revolving parts were electromagnetic. He formulated the concept of the dynamo at least 6 years before Siemens and Wheatstone but didn't patent it as he thought he wasn't the first to realise this. In essence the concept is that instead of permanent magnets, two electromagnets opposite to each other induce the magnetic field around the rotor. It was also the discovery of the principle of self-excitation.

In the years of 1831–1832, Michael Faraday discovered the operating principle of electromagnetic generators. The principle, later called Faraday's law, is that an electromotive force is generated in an electrical conductor which encircles a varying magnetic flux. He also built the first electromagnetic generator, called the Faraday disk, a type of homopolar generator, using a copper disc rotating between the poles of a horseshoe magnet. It produced a small DC voltage.

This design was inefficient, due to self-cancelling counterflows of current in regions that were not under the influence of the magnetic field. While current was induced directly underneath the magnet, the current would circulate backwards in regions that were outside the influence of the magnetic field. This counterflow limited the power output to the pickup wires, and induced waste heating of the copper disc. Later homopolar generators would solve this problem by using an array of magnets arranged around the disc perimeter to maintain a steady field effect in one current-flow direction.

Another disadvantage was that the output voltage was very low, due to the single current path through the magnetic flux. Experimenters found that using multiple turns of wire in a coil could produce higher, more useful voltages. Since the output voltage is proportional to the number of turns, generators could be easily designed to produce any desired voltage by varying the number of turns. Wire windings became a basic feature of all subsequent generator designs.

The dynamo was the first electrical generator capable of delivering power for industry. The dynamo uses electromagnetic induction to convert mechanical rotation into direct current through the use of a commutator. The first dynamo was built by Hippolyte Pixii in 1832.

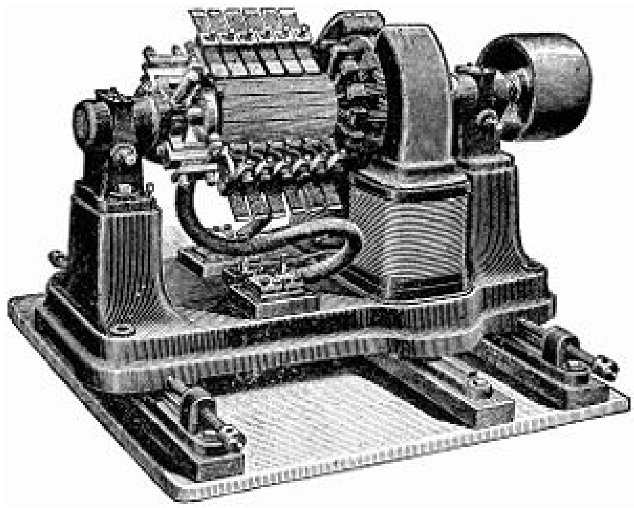

Figure: *Dynamos are no longer used for power generation due to the size and complexity of the commutator needed for high power applications. This large belt-driven high-current dynamo produced 310 amperes at 7 volts, or 2,170 watts, when spinning at 1400 RPM.*

A dynamo machine consists of a stationary structure, which provides a constant magnetic field, and a set of rotating windings which turn within that field. On small machines the constant magnetic field may be provided by one or more permanent magnets; larger machines have the constant magnetic field provided by one or more electromagnets, which are usually called field coils.

Through a series of accidental discoveries, the dynamo became the source of many later inventions, including the DC electric motor, the AC alternator, the AC synchronous motor, and the rotary converter.

Alternating current generating systems were known in simple forms from the discovery of the magnetic induction of electric current. The early machines were developed by pioneers such as Michael Faraday and Hippolyte Pixii.

Faraday developed the "rotating rectangle", whose operation was *heteropolar* - each active conductor passed successively through regions where the magnetic field was in opposite directions. The first public demonstration of a more robust "alternator system" took place in 1886. Large two-phase alternating current generators were built by a British electrician, J.E.H. Gordon, in 1882. Lord Kelvin and Sebastian Ferranti also developed early alternators, producing frequencies between 100 and 300 Hz. In 1891, Nikola Tesla patented a practical "high-frequency" alternator (which operated around 15 kHz). After 1891, polyphase alternators were introduced to supply currents of multiple differing phases.

Later alternators were designed for varying alternating-current frequencies between sixteen and about one hundred hertz, for use with arc lighting, incandescent lighting and electric motors.

Large power generation dynamos are now rarely seen due to the now nearly universal use of alternating current for power distribution. Before the adoption of AC, very large direct-current dynamos were the only means of power generation and distribution. AC has come to dominate due to the ability of AC to be easily transformed to and from very high voltages to permit low losses over large distances.

Electromagnetic Generators

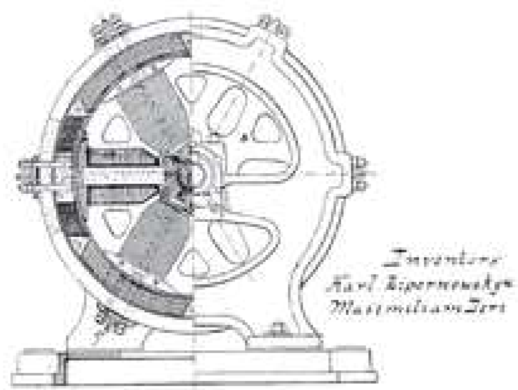

Figure: *"Dynamo Electric Machine" (end view, partly section, U.S. Patent 284,110), 1883*

Dynamo: A dynamo is an electrical generator that produces direct current with the use of a commutator. Dynamos were the first electrical generators capable of delivering power for industry, and the foundation upon which many other later electric-power conversion devices were based, including the electric motor, the alternating-current alternator, and the rotary converter.

Today, the simpler alternator dominates large scale power generation, for efficiency, reliability and cost reasons. A dynamo has the disadvantages of a mechanical commutator. Also, converting alternating to direct current using power rectification devices (vacuum tube or more recently solid state) is effective and usually economic.

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Geotechnical Investigation

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Geotechnical investigations are performed by geotechnical engineers or engineering geologists to obtain information on the physical properties of soil and rock around a site to design earthworks and foundations for proposed structures and for repair of distress to earthworks and structures caused by subsurface conditions. A geotechnical investigation will include surface exploration and subsurface exploration of a site. Sometimes, geophysical methods are used to obtain data about sites. Subsurface exploration usually involves soil sampling and laboratory tests of the soil samples retrieved. Surface exploration can include geologic mapping, geophysical methods, and photogrammetry, or it can be as simple as a geotechnical professional walking around on the site to observe the physical conditions at the site.

To obtain information about the soil conditions below the surface, some form of subsurface exploration is required. Methods of observing the soils below the surface, obtaining samples, and determining physical properties of the soils and rocks include test pits, trenching (particularly for locating faults and slide planes), boring, and in situ tests.

Soil Sampling

Borings come in two main varieties, large-diameter and small-diameter. Large-diameter borings are rarely used due to safety concerns and expense, but are sometimes used to allow a geologist or engineer to visually and manually examine the soil and rock stratigraphy in-situ. Small-diameter borings are frequently used to allow a geologist or engineer examine soil or rock cuttings or to retrieve samples at depth using soil samplers, and to perform in-place soil tests.

Soil samples are often categorised as being either “disturbed” or “undisturbed;” however, “undisturbed” samples are not truly undisturbed. A disturbed sample is one in which the structure of the soil has been changed sufficiently that tests of structural properties of the soil will not be representative of in-situ conditions, and only properties of the soil grains (e.g., grain size distribution, Atterberg limits, and possibly the water content) can be accurately determined. An undisturbed sample is one where the condition of the soil in the sample is close enough to the conditions of the soil in-situ to allow tests of structural properties of the soil to be used to approximate the properties of the soil in-situ.

Offshore soil collection introduces many difficult variables. In shallow water, work can be done off a barge. In deeper water a ship will be required. Deepwater soil samplers are normally variants of Kullenberg-type samplers, a modification on a basic gravity corer using a piston (Lunne and Long, 2006). Seabed samplers are also available, which push the collection tube slowly into the soil.

Soil Samplers

Soil samples are taken using a variety of samplers; some provide only disturbed samples, while others can provide relatively undisturbed samples.

Shovel: Samples can be obtained by digging out soil from the site. Samples taken this way are disturbed samples.

Hand/Machine Driven Auger: This sampler typically consists of a short cylinder with a cutting edge attached to a rod and handle. The sampler is advanced by a combination of rotation and downward force. Samples taken this way are disturbed samples.

Continuous Flight Auger: A method of sampling using an auger as a corkscrew. The auger is screwed into the ground then lifted out. Soil is retained on the blades of the auger and kept for testing. The soil sampled this way is considered disturbed.

Split-spoon / SPT Sampler: Utilised in the 'Standard Test Method for Standard Penetration Test (SPT) and Split-Barrel Sampling of Soils' (ASTM D 1586). This sampler is typically a 18"-30" long, 2.0" outside diameter (OD) hollow tube split in half lengthwise. A hardened metal drive shoe with a 1.375" opening is attached to the bottom end, and a one-way valve and drill rod adapter at the sampler head. It is driven into the ground with a 140-pound (64 kg) hammer falling 30". The blow counts (hammer strikes) required to advance the sampler a total of 18" are counted and reported. Generally used for non-cohesive soils, samples taken this way are considered disturbed.

Modified California Sampler: Similar in concept to the SPT sampler, the sampler barrel has a larger diameter and is usually lined with metal tubes to contain samples. Samples from the Modified California Sampler are considered disturbed due to the large area ratio of the sampler (sampler wall area/sample cross sectional area).

Shelby Tube Sampler: Utilised in the 'Standard Practice for Thin-Walled Tube Sampling of Soils for Geotechnical Purposes' (ASTM D 1587). This sampler consists of a thin-walled tube with a cutting edge at the toe. A sampler head attaches the tube to the drill rod, and contains a check valve and pressure vents. Generally used in cohesive soils, this sampler is advanced into the soil layer, generally 6" less than the length of the tube. The vacuum created by the check valve and cohesion of the sample in the tube cause the sample to be retained when the tube is withdrawn. Standard ASTM dimensions are; 2" OD, 36" long, 18 gauge thickness; 3" OD, 36" long, 16 gauge thickness; and 5" OD, 54" long, 11 gauge thickness. It should be noted that ASTM allows other diameters as long as they are proportional to the standardised tube designs, and tube length is to be suited for field conditions. Soil sampled in this manner is considered undisturbed.

Piston Samplers: These samplers are thin-walled metal tubes which contain a piston at the tip. The samplers are pushed into the bottom of a borehole, with the piston remaining at the surface of the soil while the tube slides past it. These samplers will return undisturbed samples in soft soils, but are difficult to advance in sands and stiff clays, and can be damaged (compromising the sample) if gravel is encountered. The Livingstone corer, developed by D. A. Livingstone, is a commonly used piston sampler. A modification of the Livingstone corer with a serrated coring head allows it to be rotated to cut through subsurface vegetable matter such as small roots or buried twigs.

Pitcher Barrel Sampler: This sampler is similar to piston samplers, except that there is no piston. There are pressure-relief holes near the top of the sampler to prevent pressure buildup of water or air above the soil sample.

In-situ Tests

A standard penetration test (SPT) is an in-situ dynamic penetration test designed to provide information on the properties of soil, while also collecting a disturbed soil sample for grain-size analysis and soil classification. A cone penetration test (CPT) is performed using an instrumented probe with a conical tip, pushed into the soil hydraulically at a constant rate. A basic CPT instrument reports tip resistance and shear resistance along the cylindrical barrel. CPT data has been correlated to soil properties. Sometimes instruments other than the basic CPT probe are used, including:

CPTu - Piezocone Penetrometer: This probe is advanced using the same equipment as a regular CPT probe, but the probe has an additional instrument which measures the groundwater pressure as the probe is advanced.

SCPTu - Seismic Piezocone Penetrometer: This probe is advanced using the same equipment as a CPT or CPTu probe, but the probe is also equipped with either geophones or accelerometers to detect shear waves and/or pressure waves produced by a source at the surface.

Full Flow Penetrometers - T-bar, Ball, and Plate: These probes are used in extremely soft clay soils (such as sea-floor deposits) and are advanced in the same manner as the CPT. As their names imply, the T-bar is a cylindrical bar attached at right angles to the drill string forming what look likes a T, the ball is a large sphere, and the plate is flat circular plate. In soft clays, soil flows around the probe similar to a viscous fluid.

The pressure due to overburden stress and pore water pressure is equal on all sides of the probes (unlike with CPT's), so no correction is necessary, reducing a source of error and increasing accuracy. Especially desired in soft soils due to the very low loads on the measuring sensors. Full flow probes can also be cycled up and down to measure remolded soil resistance. Ultimately the geotechnical professional can use the measured penetration resistance to estimate undrained and remolded shear strengths.

Flat Plate Dilatometer Test (DMT) is a flat plate probe often advanced using CPT rigs, but can also be advanced from conventional drill rigs. A diaphragm on the plate applies a lateral force to the soil materials and measures the strain induced for various levels of applied stress at the desired depth interval.

Laboratory Tests

A wide variety of laboratory tests can be performed on soils to measure a wide variety of soil properties. Some soil properties are intrinsic to the composition of the soil matrix and are not affected by sample disturbance, while other properties depend on the structure of the soil as well as its composition, and can only be effectively tested on relatively undisturbed samples. Some soil tests measure direct properties of the soil, while others measure "index properties" which provide useful information about the soil without directly measuring the property desired.

Atterberg Limits

The Atterberg limits define the boundaries of several states of consistency for plastic soils. The boundaries are defined by the amount of water a soil needs to be at one of those boundaries. The boundaries are called the plastic limit and the liquid limit, and the difference between them is called the plasticity index. The shrinkage limit is also a part of the Atterberg limits. The results of this test can be used to help predict other engineering properties.

California Bearing Ratio

ASTM D 1883. A test to determine the aptitude of a soil or aggregate sample as a road subgrade. A plunger is pushed into a compacted sample, and its resistance is measured. This test was developed by Caltrans, but it is no longer used in the Caltrans pavement design method. It is still used as a cheap method to estimate the resilient modulus.

Direct Shear Test

ASTM D3080. The direct shear test determines the consolidated, drained strength properties of a sample. A constant strain rate is applied to a single shear plane under a normal load, and the load response is measured. If this test is performed with different normal loads, the common shear strength parameters can be determined.

Expansion Index Test

This test uses a remolded soil sample to determine the Expansion Index (EI), an empirical value required by building design codes, at a water content of 50% for expansive soils, like expansive clays.

Hydraulic Conductivity Tests

There are several tests available to determine a soil's hydraulic conductivity. They include the constant head, falling head, and constant flow methods. The soil samples tested can be any type include remolded, undisturbed, and compacted samples.

Oedometer Test

This can be used to determine consolidation (ASTM D2435) and swelling (ASTM D4546) parameters.

Particle-size Analysis

This is done to determine the soil gradation. Coarser particles are separated in the sieve analysis portion, and the finer particles are analysed with a hydrometer. The distinction between coarse and fine particles is usually made at 75 ìm. The sieve analysis shakes the sample through progressively smaller meshes to determine its gradation. The hydrometer analysis uses the rate of sedimentation to determine particle gradation.

R-Value Test

California Test 301 This test measures the lateral response of a compacted sample of soil or aggregate to a vertically applied pressure under specific conditions. This test is used by Caltrans for pavement design, replacing the California bearing ratio test.

Soil Compaction Tests

Standard Proctor (ASTM D698), Modified Proctor (ASTM D1557), and California Test 216. These tests are used to determine the maximum unit weight and optimal water content a soil can achieve for a given compaction effort.

Soil Suction Tests

ASTM D5298.

Triaxial Shear Tests

This is a type of test that is used to determine the shear strength properties of a soil. It can simulate the confining pressure a soil would see deep into the ground. It can also simulate drained and undrained conditions.

Unconfined Compression Test

ASTM D698. This test compresses a soil sample to measure its strength. The modifier "unconfined" contrasts this test to the triaxial shear test.

Water Content

This test provides the water content of the soil, normally expressed as a percentage of the weight of water to the dry weight of the soil.

Geophysical Exploration

Geophysical methods are used in geotechnical investigations to evaluate a site's behaviour in a seismic event. By measuring a soil's shear wave velocity, the dynamic response of that soil can be estimated. There are a number of methods used to determine a site's shear wave velocity:

- Crosshole method
- Downhole method (with a seismic CPT or a substitute device)
- Surface wave reflection or refraction
- Suspension logging (also known as P-S logging or Oyo logging)
- Spectral analysis of surface waves (SASW)
- Modal Analysis of Surface waves (MASW)
- Reflection microtremor (ReMi).

Ambient Vibrations

Various types of vibration sources are always producing so called Ambient Vibrations on the Earth ground (also called ambient noise). These vibrations are mostly surface waves (Rayleigh waves, Love waves) propagating on the surface. Low frequency waves (below 1 Hz) are generally called microseisms and high frequency waves (above 1 Hz) are called microtremors. These ambient vibrations are used in practice to derive the elastic properties of the ground and the low-strain dynamic properties of civil-engineering structures (bridges, buildings, dams...). This information is useful for different purposes : fundamental seismology, engineering seismology, Earthquake engineering, Seismic microzonation, Structural health monitoring, but also Hydrology, Geotechnical Engineering, etc.

Physical Origin of the Ambient Vibrations

Bonnefoy-Claudet et al. reviewed the scientific work studying the origin of the noise wavefield. At low frequency (below 1 Hz), the noise sources are natural and mostly due to ocean waves. In particular the peak between 0.1 and 0.3 Hz is clearly associated with the interaction of water waves of nearly equal frequencies but opposite directions. At high frequency (above 1 Hz), the wavefield is mainly produced by human activities (road traffic, industrial work...) but there are also natural sources like rivers. Around 1 Hz, the local atmospheric conditions (wind...) are also a major source of ground vibrations.

The amplitude of ground ambient vibrations is typically in the range of 1e-6 m, i.e. in the order of the tenth of micrometers to tens of micrometers. Peterson provided high and low noise models as a function of frequency. The ambient wave field is made of a small amount of body waves (P- and S-waves), and a most generally predominant part ofsurface waves, i.e. Love and Rayleigh waves. Theses waves are dispersive, i.e. their phase velocity varies with frequency (most generally, it decreases with increasing frequency). The dispersion curve (phase velocity or slowness as a function of frequency) is tightly related to the variations of the shear-wave velocity with depth in the different ground layers: it can thus be used as a non-invasive tool to investigate the underground structure.

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Biodiversity and its Conservation

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Biodiversity

Biodiversity is the degree of variation of life forms within a given species, ecosystem, biome, or planet. Terrestrial biodiversity tends to be highest at low latitudes near the equator, which seems to be the result of the warm climate and high primary productivity. Marine biodiversity tends to be highest along coasts in the Western Pacific, where sea surface temperature is highest and in mid-latitudinal band in all oceans. Biodiversity generally tends to cluster in hotspots, and has been increasing through time but will be likely to slow in the future.

Rapid environmental changes typically cause mass extinctions. One estimate is that <1%-3% of the species that have existed on Earth are extant.

Since life began on Earth, five major mass extinctions and several minor events have led to large and sudden drops in biodiversity. The Phanerozoic eon (the last 540 million years) marked a rapid growth in biodiversity via the Cambrian explosion—a period during which the majority of multicellular phyla first appeared. The next 400 million years included repeated, massive biodiversity losses classified as mass extinction events. In the Carboniferous, rainforest collapse led to a great loss of plant and animal life. The Permian–Triassic extinction event, 251 million years ago, was the worst; vertebrate recovery took 30 million years. The most recent, the Cretaceous–Paleogene extinction event, occurred 65 million years ago and has often attracted more attention than others because it resulted in the extinction of the dinosaurs. The period since the emergence of humans has displayed an ongoing biodiversity reduction and an accompanying loss of genetic diversity. Named the Holocene extinction, the reduction is caused primarily by human impacts, particularly habitat destruction. Conversely, biodiversity impacts human health in a number of ways, both positively and negatively.

The United Nations designated 2011-2020 as the United Nations Decade on Biodiversity.

Etymology

The term biological diversity was used first by wildlife scientist and conservationist Raymond F. Dasmann in the 1968 lay book *A Different Kind of Country* advocating conservation. The term was widely adopted only after more than a decade, when in the 1980s it came into common usage in science and environmental policy. Thomas Lovejoy, in the foreword to the book *Conservation Biology*, introduced the term to the scientific community. Until then the term "natural diversity" was common, introduced by The Science Division of The Nature Conservancy in an important 1975 study, "The Preservation of Natural Diversity." By the early 1980s TNC's Science programme and its head, Robert E. Jenkins, Lovejoy and other leading conservation scientists at the time in America advocated the use of the term "biological diversity".

The term's contracted form *biodiversity* may have been coined by W.G. Rosen in 1985 while planning the 1986 *National Forum on Biological Diversity* organized by the National Research Council (NRC). It first appeared in a publication in 1988 when sociobiologist E. O. Wilson used it as the title of the proceedings of that forum.

Since this period the term has achieved widespread use among biologists, environmentalists, political leaders, and concerned citizens. A similar term in the United States is "natural heritage." It predates the others and is more accepted by the wider audience interested in conservation. Broader than biodiversity, it includes geology and landforms.

Definitions

"Biodiversity" is most commonly used to replace the more clearly defined and long established terms, species diversity and species richness. Biologists most often define biodiversity as the "totality of genes, species, and ecosystems of a region". An advantage of this definition is that it seems to describe most circumstances and presents a unified view of the traditional three levels at which biological variety has been identified:

- species diversity
- ecosystem diversity
- genetic diversity

In 2003 Professor Anthony Campbell at Cardiff University, UK and the Darwin Centre, Pembrokeshire, defined a fourth level: Molecular Diversity.

This multilevel construct is consistent with Dasmann and Lovejoy. An explicit definition consistent with this interpretation was first given in a paper by Bruce A. Wilcox commissioned by the International Union for the Conservation of Nature and Natural Resources (IUCN) for the 1982 World National Parks Conference. Wilcox's definition was "Biological diversity is the variety of life forms...at all levels of biological systems (i.e., molecular, organismic, population, species and ecosystem)...". The 1992 United Nations Earth Summit defined "biological diversity" as "the variability among living organisms from all sources, including, 'inter alia', terrestrial, marine, and other aquatic ecosystems, and the ecological complexes of which they are part: this includes diversity within species, between species and of ecosystems". This definition is used in the United Nations Convention on Biological Diversity.

One textbook's definition is "variation of life at all levels of biological organization".

Genetically biodiversity can be defined as the diversity of alleles, genes, and organisms. They study processes such as mutation and gene transfer that drive evolution.

Measuring diversity at one level in a group of organisms may not precisely correspond to diversity at other levels. However, tetrapod (terrestrial vertebrates) taxonomic and ecological diversity shows a very close correlation.

Distribution

Biodiversity is not evenly distributed, rather it varies greatly across the globe as well as within regions. Among other factors, the diversity of all living things (biota) depends on temperature, precipitation, altitude, soils, geography and the presence of other species. The study of the spatial distribution of organisms, species, and ecosystems, is the science of biogeography.

Diversity consistently measures higher in the tropics and in other localized regions such as the Cape Floristic Region and lower in polar regions generally. Rain forests that have had wet climates for a long time, such as Yasuni National Park in Ecuador, have particularly high biodiversity. Terrestrial biodiversity is up to 25 times greater than ocean biodiversity. Although a recent discovered method put the total number of species on Earth at 8.7 million of which 2.1 million were estimated to live in the ocean, however this estimate seems to under-represent diversity of microorganisms.

Latitudinal Gradients

Generally, there is an increase in biodiversity from the poles to the tropics. Thus localities at lower latitudes have more species than localities at higher latitudes. This is often referred to as the latitudinal gradient in species diversity. Several ecological mechanisms may contribute to the gradient, but the ultimate factor behind many of them is the greater mean temperature at the equator compared to that of the poles. Even though terrestrial biodiversity declines from the equator to the poles, some studies claim that this characteristic is unverified in aquatic ecosystems, especially in marine ecosystems. The latitudinal distribution of parasites does not follow this rule.

Hotspots

A biodiversity hotspot is a region with a high level of endemic species that is under threat from humans. The term hotspot was introduced in 1988 by Dr. Sabina Virk. While hotspots are spread all over the world, the majority are forest areas and most are located in the tropics.

Brazil's Atlantic Forest is considered one such hotspot, containing roughly 20,000 plant species, 1,350 vertebrates, and millions of insects, about half of which occur nowhere else. The island of Madagascar, particularly the unique Madagascar dry deciduous forests and lowland rainforests, possess a high ratio of endemism. Since the island separated from mainland Africa 65 million years ago, many species and ecosystems have evolved independently. Indonesia's 17,000 islands cover 735,355 square miles (1,904,560 km^2) contain 10% of the world's flowering plants, 12% of mammals and 17% of reptiles, amphibians and birds—along with nearly 240 million people. Many regions of high biodiversity and/or endemism arise from specialized habitats which require unusual adaptations, for example alpine environments in high mountains, or Northern European peat bogs.

Accurately measuring differences in biodiversity can be difficult. Selection bias amongst researchers may contribute to biased empirical research for modern estimates of biodiversity. In 1768 Rev. Gilbert White succinctly observed of his Selborne, Hampshire "all nature is so full, that district produces the most variety which is the most examined."

Evolution

Biodiversity is the result of 3.5 billion years of evolution. The origin of life has not been definitely established by science, however some evidence suggests that life may already have been well-established only a few hundred million years after the formation of the Earth. Until approximately 600 million years ago, all life consisted of archaea, bacteria, protozoans and similar single-celled organisms.

The history of biodiversity during the Phanerozoic (the last 540 million years), starts with rapid growth during the Cambrian explosion—a period during which nearly every phylum of multicellular organisms first appeared. Over the next 400 million years or so, invertebrate diversity showed little overall trend, and vertebrate diversity shows an overall exponential trend. This dramatic rise in diversity was marked by periodic, massive losses of diversity classified as mass extinction events. A significant loss occurred when rainforests collapsed in the carboniferous. The worst was the Permo-Triassic extinction, 251 million years ago. Vertebrates took 30 million years to recover from this event.

The fossil record suggests that the last few million years featured the greatest biodiversity in history. However, not all scientists support this view, since there is uncertainty as to how strongly the fossil record is biased by the greater availability and preservation of recent geologic sections.

Some scientists believe that corrected for sampling artifacts, modern biodiversity may not be much different from biodiversity 300 million years ago., whereas others consider the fossil record reasonably reflective of the diversification of life. Estimates of the present global macroscopic species diversity vary from 2 million to 100 million, with a best estimate of somewhere near 9 million, the vast majority arthropods. Diversity appears to increase continually in the absence of natural selection.

Evolutionary Diversification

The existence of a "global carrying capacity", limiting the amount of life that can live at once, is debated, as is the question of whether such a limit would also cap the number of species. While records of life in the sea shows a logistic pattern of growth, life on land (insects, plants and tetrapods) shows an exponential rise in diversity. As one author states, "Tetrapods have not yet invaded 64 per cent of potentially habitable modes, and it could be that without human influence the ecological and taxonomic diversity of tetrapods would continue to increase in an exponential fashion until most or all of the available ecospace is filled."

On the other hand, changes through the Phanerozoic correlate much better with the hyperbolic model (widely used in population biology, demography and macrosociology, as well as fossil biodiversity) than with exponential and logistic models.

The latter models imply that changes in diversity are guided by a first-order positive feedback (more ancestors, more descendants) and/or a negative feedback arising from resource limitation. Hyperbolic model implies a second-order positive feedback. The hyperbolic pattern of the world population growth arises from a second-order positive feedback between the population size and the rate of technological growth. The hyperbolic character of biodiversity growth can be similarly accounted for by a feedback between diversity and community structure complexity. The similarity between the curves of biodiversity and human population probably comes from the fact that both are derived from the interference of the hyperbolic trend with cyclical and stochastic dynamics.

Most biologists agree however that the period since human emergence is part of a new mass extinction, named the Holocene extinction event, caused primarily by the impact humans are having on the environment. It has been argued that the present rate of extinction is sufficient to eliminate most species on the planet Earth within 100 years.

New species are regularly discovered (on average between 5–10,000 new species each year, most of them insects) and many, though discovered, are not yet classified (estimates are that nearly 90% of all arthropods are not yet classified). Most of the terrestrial diversity is found in tropical forests and in general, land has more species than the ocean; some 8.7 million species may exists on Earth, of which some 2.1 million live in the ocean.

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Population Dynamics

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Ecotype

In evolutionary ecology, an ecotype, sometimes called ecospecies, describes a genetically distinct geographic variety, population or race within species (or among closely related), which is adapted to specific environmental conditions. Typically, ecotypes exhibit phenotypic differences (such as in morphology or physiology) stemming from environmental heterogeneity and are capable of interbreeding with other geographically adjacent ecotypes without loss of fertility or vigor.

Range and Distribution

Experiments indicate that sometimes ecotypes manifest only when separated by great spatial distances (of the order of 1000 km). This is due to hybridization whereby different but adjacent varieties of the same species (or generally of the same taxonomic rank) interbreed, thus overcoming local selection. However other studies reveal that the opposite may happen, i.e. ecotypes revealing at very small scales (of the order of 10 m), within populations, and despite hybridization.

In ecotypes, it is common for continuous, gradual geographic variation to impose analogous phenotypic and/or genetic variation. This situation is called cline. A well-known example of cline is the skin colour gradation in indigenous human populations worldwide, which is related to latitude and amounts of sunlight. But often the distribution of ecotypes is bimodal or multimodal. This means that ecotypes may display two or more distinct and discontinuous phenotypes even within the same population. Such phenomenon may lead to speciation and can occur if conditions in a local environment change dramatically through space or time.

Examples

- Earthworms fall into four different ecotypes. Compost earthworm prefer warm and moist environments with a ready supply of fresh compost material. Epigeic earthworms live on the surface of the soil in leaf litter and tend not to make burrows but live in and feed on the leaf litter. Endogeic earthworms live in and feed on the soil, making horizontal burrows through the soil to move around and to feed and they will reuse these burrows to a certain extent. Anecic earthworms make permanent vertical burrows in soil, feeding on leaves on the soil surface that they drag into their burrows.

- Tundra reindeer and forest (or woodland) reindeer are two ecotypes of reindeer. The first migrate (travelling 5,000 km) annually between the two environments in large numbers whereas the other (who are much fewer) remain in the forest for the summer. Currently, and since 1961 classification, tundra reindeer comprise five subspecies and woodland reindeer two.
- *Arabis fecunda*, a herb endemic to some calcareous soils of Montana, USA, can be divided into two ecotypes. The one "low elevation" group lives near the ground in an arid, warm environment and has thus developed a significantly greater tolerance against drought than the "high elevation" group. The two ecotypes are separated by a horizontal distance of about 100 km.
- It is commonly accepted that the Tucuxi dolphin has two ecotypes - the riverine ecotype found in some South American rivers and the pelagic ecotype found in the South Atlantic Ocean. Similarly, it is accepted that the Common Bottlenose Dolphin has two ecotypes in the Western North Atlantic.
- The Warbler finch and the Cocos Island Finch are viewed as separate ecotypes.
- The Scots Pine (*Pinus sylvestris*) has 20 different ecotypes in an area from Scotland to Siberia, all capable of interbreeding.
- A very subtle case of ecotype is the following: It has been observed that two populations of the same *Helix* snail species separated by only a few hundred kilometers prefer not to cross-mate, i.e. they reject one another as mates. This event probably occurs during the process of courtship, which may last for hours.

Terminology

Ecotypes have no main taxonomic rank in modern biological classification. However some scientists consider them "taxonomically equivalent to subspecies". This is true in the sense that ecotypes can be sometimes classified as subspecies and the opposite.

Ecotypes are closely related to morphs. In the context of evolutionary biology, genetic polymorphism is the occurrence in equilibrium of two or more distinctly different phenotypes within a population of a species, in other words, the occurrence of more than one form or morph. The frequency of these discontinuous forms (even that of the rarest) is too high to be explained by mutation.

In order to be classified as such, morphs must occupy the same habitat at the same time and belong to a panmictic population (whose all members can potentially interbreed). Polymorphism is actively and steadily maintained in populations of species by natural selection (most famously sexual dimorphism in humans) in contrast to *transient polymorphisms* where conditions in a habitat change in such a way that a "form" is being replaced completely by another.

In fact, Begon, Townsend and Harper assert that

> *There is not always clear distinction between local ecotypes and genetic polymorphisms.*

The notions "form" and "ecotype" may appear to correspond to a static phenomenon, however this is not always the case. Evolution occurs continuously both in time and space, so that two ecotypes or forms may qualify as distinct species in only a few generations. Begon, Townsend and Harper use an illuminating analogy on this:

> *… the origin of a species, whether allopatric or sympatric, is a process, not an event. For the formation of a new species, like the boiling of an egg, there is some freedom to argue about when it is completed.*

Thus ecotypes and morphs can be thought of as precursory steps of potential speciation.

Population Ecology

Population ecology is a sub-field of ecology that deals with the dynamics of speciespopulations and how these populations interact with the environment. It is the study of how the population sizes of species living together in groups change over time and space.

The development of population ecology owes much to demography and actuariallife tables. Population ecology is important in conservation biology, especially in the development of population viability analysis (PVA) which makes it possible to predict the long-term probability of a species persisting in a given habitat patch, such as a national park. Although population ecology is a subfield of biology, it provides interesting problems for mathematicians and statisticians who work in population dynamics.

Metapopulation

Populations are also studied and conceptualized through the "metapopulation" concept. The metapopulation concept was introduced in 1969:

> *"as a population of populations which go extinct locally and recolonize."*

Metapopulation ecology is a simplified model of the landscape into patches of varying levels of quality.[13] Patches are either occupied or they are not. Migrants moving among the patches are structured into metapopulations either as sources or sinks. Source patches are productive sites that generate a seasonal supply of migrants to other patch locations. Sink patches are unproductive sites that only receive migrants. In metapopulation terminology there are emigrants (individuals that leave a patch) and immigrants (individuals that move into a patch). Metapopulation models examine patch dynamics over time to answer questions about spatial and demographic ecology. An important concept in metapopulation ecology is the rescue effect, where small patches of lower quality (i.e., sinks) are maintained by a seasonal influx of new immigrants. Metapopulation structure evolves from year to year, where some patches are sinks, such as dry years, and become sources when conditions are more favorable. Ecologists utilize a mixture of computer models and field studies to explain metapopulation structure.

History

The older term, autecology, refers to roughly the same field of study as population ecology. It derives from the division of ecology into autecology—the study of individual species in relation to the environment—and synecology—the study of groups of organisms in relation to the environment—or community ecology. Odum (1959, p. 8) considered that synecology should be divided into population ecology, community ecology, and ecosystem ecology, defining autecology as essentially "species ecology." However, for some time biologists have recognized that the more significant level of organization of a species is a population, because at this level the species gene pool is most coherent. In fact, Odum regarded "autecology" as no longer a "present tendency" in ecology (i.e., an archaic term), although included "species ecology"—studies emphasizing life history and behavior as adaptations to the environment of individual organisms or species—as one of four subdivisions of ecology.

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Degeneracy: Stalling and Cycling

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If the values of all basic variables are strictly positive, then a pivot must result in an improvement in the objective value. When this is always the case no set of basic variables occurs twice and the simplex algorithm must terminate after a finite number of steps. Basic feasible solutions where at least one of the *basic* variables is zero are called *degenerate* and may result in pivots for which there is no improvement in the objective value. In this case there is no actual change in the solution but only a change in the set of basic variables. When several such pivots occur in succession, there is no improvement; in large industrial applications, degeneracy is common and such "*stalling*" is notable. Worse than stalling is the possibility the same set of basic variables occurs twice, in which case, the deterministic pivoting rules of the simplex algorithm will produce an infinite loop, or "cycl≥. While degeneracy is the rule in practice and stalling is common, cycling is rare in practice. A discussion of an example of practical cycling occurs in Padberg. Bland's rule prevents cycling and thus guarantee that the simplex algorithm always terminates. Another pivoting algorithm, the criss-cross algorithm never cycles on linear programmes.

Efficiency

The simplex method is remarkably efficient in practice and was a great improvement over earlier methods such as Fourier–Motzkin elimination. However, in 1972, Klee and Minty gave an example showing that the worst-case complexity of simplex method as formulated by Dantzig is exponential time. Since then, for almost every variation on the method, it has been shown that there is a family of linear programmes for which it performs badly. It is an open question if there is a variation with polynomial time, or even sub-exponential worst-case complexity.

Analyzing and quantifying the observation that the simplex algorithm is efficient in practice, even though it has exponential worst-case complexity, has led to the development of other measures of complexity. The simplex algorithm has polynomial-time average-case complexity under various probability distributions, with the precise average-case performance of the simplex algorithm depending on the choice of a probability distribution for the random matrices. Another approach to studying "typical phenoma" uses Baire category theory from general topology, and to show that (topologically) "most" matrices can be solved by the simplex algorithm in a polynomial number of steps.

Another method to analyze the performance of the simplex algorithm studies the behaviour of worst-case scenarios under small perturbation – are worst-case scenarios stable under a small change (in the sense of structural stability), or do they become tractable? Formally, this method uses random problems to which is added a Gaussian random vector ("smoothed complexity").

Other Algorithms

Other algorithms for solving linear-programming problems are described in the linear-programming article. Another basis-exchange pivoting algorithm is the criss-cross algorithm. There are polynomial-time algorithms for linear programming that use interior point methods: These include Khachiyan's ellipsoidal algorithm, Karmarkar's projective algorithm, and path-following algorithms.

Linear-fractional Programming

In mathematical optimization, linear-fractional programming (LFP) is a generalization of linear programming (LP). Whereas the objective function in linear programmes are linear functions, the objective function in a linear-fractional programme is a ratio of two linear functions. A linear programme can be regarded as a special case of a linear-fractional programme in which the denominator is the constant function one.

Relation to Linear Programming

Both linear programming and linear-fractional programming represent optimization problems using linear equations and linear inequalities, which for each problem-instance define a feasible set. Fractional linear programmes have a richer set of objective functions. Informally, linear programming computes a policy delivering the best outcome, such as maximum profit or lowest cost. In contrast, a linear-fractional programming is used to achieve the highest *ratio* of outcome to cost, the ratio representing the highest efficiency. For example, in the context of LP we maximize the objective function profit = income " cost and might obtain maximal profit of \$100 (= \$1100 of income " \$1000 of cost). Thus, in LP we have an efficiency of \$100/\$1000 = 0.1. Using LFP we might obtain an efficiency of \$10/\$50 = 0.2 with a profit of only \$10, which requires only \$50 of investment however.

Definition

Formally, a linear-fractional programme is defined as the problem of maximizing (or minimizing) a ratio of affine functions over a polyhedron,

$$\begin{array}{ll} \max imize & \dfrac{c^T x + \alpha}{d^T x + \beta} \\ subject\ to & Ax \le b, \end{array}$$

where $x \in \mathbb{R}^n$ represents the vector of variables to be determined, $c,d \in \mathbb{R}^n$ and $b \in \mathbb{R}^m$ are vectors of (known) coefficients, $A \in \mathbb{R}^{m\times n}$ is a (known) matrix of coefficients and $\alpha,\beta \in \mathbb{R}$ are constants.

The constraints have to restrict the feasible region to $\{x \mid d^T x + \beta > 0\}$, i.e. the region on which the denominator is positive. Alternatively, the denominator of the objective function has to be strictly negative in the entire feasible region.

Transformation to a Linear Programme

Under the assumption that the feasible region is non-empty and bounded, the Charnes-Cooper transformation $y = \dfrac{1}{d^T x + \beta} \cdot x; t = \dfrac{1}{d^T x + \beta}$ translates the linear-fractional programme above to the equivalent linear programme

$$\begin{array}{ll} \max imize & c^T y + \alpha t \\ subject\ to & Ay \le bt \\ & d^T y + \beta t = 1 \\ & t \ge 0. \end{array}$$

Duality

Let the dual variables associated with the constraints $Ay - bt \le 0$ and $\mathrm{d}^T \mathrm{y} + \beta t - 1 = 0$ be denoted by u and λ, respectively. Then the dual of the LFP above is

$$\begin{array}{ll} \min imize & \lambda \\ subject\ to & A^T u + \lambda d = c \\ & -b^T u + \lambda\beta \ge \alpha \\ & u \in \mathbb{R}^n_+, \lambda \in \mathbb{R}, \end{array}$$

which is an LP and which coincides with the dual of the equivalent linear programme resulting from the Charnes-Cooper transformation.

Properties of and Algorithms for linear-fractional Programmes

The objective function in a linear-fractional problem is both quasiconcave and quasiconvex (hence quasilinear) with a monotone property, pseudoconvexity, which is a stronger property than quasiconvexity. A linear-fractional objective function is both pseudoconvex and pseudoconcave, hence pseudolinear. Since an LFP can be transformed to an LP, it can be solved using any LP solution method, such as the simplex algorithm (of George B. Dantzig), the criss-cross algorithm, or interior-point methods.

Criss-cross Algorithm

Like the simplex algorithm of Dantzig, the criss-cross algorithm is a basis-exchange algorithm that pivots between bases. However, the criss-cross algorithm need not maintain feasibility, but can pivot rather from a feasible basis to an infeasible basis.

The criss-cross algorithm does not have polynomial time-complexity for linear programming. Both algorithms visit all 2^D corners of a (perturbed) cube in dimension D, the Klee–Minty cube, in the worst case.

Conic Sampling Algorithm of Serang

Like other basis-exchange algorithms, Serang's conic sampling algorithm moves between vertices; but where the simplex algorithm moves along edges by removing and adding one basis at a time, the conic sampling method exchanges multiple bases at a time, and is not restricted to moving along edges of the polytope.

Starting at a current vertex, the conic sampling method chooses a random vector that improves the objective value without violating any adjacent constraints. The algorithm then travels along this vector until a limiting constraint is encountered. From this point, the algorithm projects the objective vector orthogonal to this limiting constraint, and moves along this orthogonal projection until a new constraint is reached.

This advancement and projection is repeated until a vertex is reached. Then, a new random vector is chosen. This process is repeated until no vector exists that can improve the objective without violating any local constraints, implying optimality. Essentially, the conic sampling method can be thought of as a vertex sampling method that randomly samples from the collection of vertices with improved objective value. If the vertices with superior objective value are sampled in a roughly uniform manner, then the expected runtime is logarithmic in the number of vertices (and thus polynomial). Sampling the vertices in this manner can permit large, beneficial jumps through the interior, and yield a substantial runtime improvent over the simplex method, especially when the number of constraints, and thus the number of potential vertices, is large; however, the tightest existing upper bound on the worst-case complexity of the conic sampling method is still exponential.

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An Introduction to Forestry Agroforestry

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Forestry is the science, art, and craft of creating, managing, using, conserving, and repairing forests and associated resources to meet desired goals, needs, and values for human benefit. Forestry is practised in plantations and natural stands. The main goal of forestry is to create and implement systems that manage forests to provide environmental supplies and services. The challenge of forestry is to create systems that are socially accepted while sustaining the resource and any other resources that might be affected.

Silviculture is a process for creating, maintaining, or restoring an appropriate balance of essential components, structures, and functions that ensure long-term ecosystem vitality, stability and resiliency (Nyland, 2007). This is done at the ground level which can contain many varieties of trees. Modern forestry generally embraces a broad range of concerns, including ecosystem services by assisting forests to provide timber as raw material for wood products, wildlife habitat, natural water quality management, recreation, landscape and community protection, employment, aesthetically appealing landscapes, biodiversity management, watershed management, erosion control, and preserving forests as 'sinks' for atmospheric carbon dioxide. A practitioner of forestry is known as a forester.

Forest ecosystems have come to be seen as the most important component of the biosphere, and forestry has emerged as a vital applied science, craft, and technology.

History

In the 5th century, monks in the then Byzantine Romagna on the Adriatic coast, established a plantation of stone pine to provide fuelwood and food. This was the beginning of the massive forest mentioned by Dante

Alighieri in his 1308 poem Divine Comedy. Formal forestry practices were developed by the Visigoths in the 7th century when, faced with the ever increasing shortage of wood, they instituted a code concerned with the preservation of oak and pine forests.

The use and management of many forest resources has a long history in China, dating from the Han Dynasty and taking place under the landowning gentry. It was also later written of by the Ming Dynasty Chinese scholar Xu Guangqi (1562–1633). In Europe, control of the land included hunting rights, and though peasants in many places were permitted to gather firewood and building timber and to graze animals, hunting rights were retained by the members of the nobility. Systematic management of forests for a sustainable yield of timber is said to have begun in the German states in the 14th century, e.g. in Nuremberg, and in 16th-century Japan. Typically, a forest was divided into specific sections and mapped; the harvest of timber was planned with an eye to regeneration.

The practice of establishing tree plantations in the British Isles was promoted by John Evelyn, though it had already acquired some popularity. Louis XIV's minister Jean-Baptiste Colbert's oak Forest of Tronçais, planted for the future use of the French Navy, matured as expected in the mid-19th century: "Colbert had thought of everything except the steamship," Fernand Braudel observed. Schools of forestry were established beginning in the late 18th century in Hesse, Russia, Austria-Hungary, Sweden, France and elsewhere in Europe. During the late 19th and early 20th centuries, forest preservation programmes were established in British India, the United States, and Europe. Many foresters were either from continental Europe (like Sir Dietrich Brandis), or educated there (like Gifford Pinchot).

The enactment and evolution of forest laws and binding regulations occurred in most Western nations in the 20th century in response to growing conservation concerns and the increasing technological capacity of logging companies.

Tropical forestry is a separate branch of forestry which deals mainly with equatorial forests that yield woods such as teak and mahogany. Sir Dietrich Brandis is considered the father of tropical forestry.

As a Science

Over the past centuries, forestry was regarded as a separate science. With the rise of ecology and environmental science, there has been a reordering in the applied sciences. In line with this view, forestry is one of three primary land-use sciences. The other two are agriculture and agroforestry. Under these headings, the fundamentals behind the management of natural forests comes by way of natural ecology. Forests or tree plantations, those whose primary purpose is the extraction of forest products, are planned and managed utilizing a mix of ecological and agroecological principles.

Today

Today a strong body of research exists regarding the management of forest ecosystems and genetic improvement of tree species and varieties. Forestry also includes the development of better methods for the planting, protecting, thinning, controlled burning, felling, extracting, and processing of timber. One of the applications of modern forestry is reforestation, in which trees are planted and tended in a given area.

Trees provide numerous environmental, social and economic benefits for people. In many regions the forest industry is of major ecological, economic, and social importance. Third-party certification systems that provide independent verification of sound forest stewardship and sustainable forestry have become commonplace in many areas since the 1990s. These certification systems were developed as a response to criticism of some forestry practices, particularly deforestation in less developed regions along with concerns over resource management in the developed world. Some certification systems are criticised for primarily acting as marketing tools and lacking in their claimed independence.

In topographically severe forested terrain, proper forestry is important for the prevention or minimization of serious soil erosion or even landslides. In areas with a high potential for landslides, forests can stabilize soils and prevent property damage or loss, human injury, or loss of life.

Public perception of forest management has become controversial, with growing public concern over perceived mismanagement of the forest and increasing demands that forest land be managed for uses other than pure timber production, for example, indigenous rights, recreation, watershed management, and preservation of wilderness, waterways and wildlife habitat. Sharp disagreements over the role of forest fires, logging, motorized recreation and other issues drives debate while the public demand for wood products continues to increase.

Foresters

Foresters work for the timber industry, government agencies, conservation groups, local authorities, urban parks boards, citizens' associations, and private landowners. The forestry profession includes a wide diversity of jobs, with educational requirements ranging from college bachelor's degrees to PhDs for highly specialized work. Industrial foresters plan forest regeneration starting with careful harvesting. Urban foresters manage trees in urban green spaces.

Foresters work in tree nurseries growing seedlings for woodland creation or regeneration projects. Foresters improve tree genetics. Forest engineers develop new building systems. Professional foresters measure and model the growth of forests with tools like geographic information systems.

Foresters may combat insect infestation, disease, forest and grassland wildfire, but increasingly allow these natural aspects of forest ecosystems to run their course when the likelihood of epidemics or risk of life or property are low. Increasingly, foresters participate in wildlife conservation planning and watershed protection. Foresters have been mainly concerned with timber management, especially reforestation, maintaining forests at prime conditions, and fire control.

Forestry plans

Foresters develop and implement forest management plans relying on mapped resource inventories showing an area's topographical features as well as its distribution of trees (by species) and other plant cover. Plans also include landowner objectives, roads, culverts, proximity to human habitation, water features and hydrological conditions, and soils information. Forest management plans typically include recommended silvicultural treatments and a timetable for their implementation.

Forest management plans include recommendations to achieve the landowner's objectives and desired future condition for the property subject to ecological, financial, logistical (e.g. access to resources), and other constraints. On some properties, plans focus on producing quality wood products for processing or sale. Hence, tree species, quantity, and form, all central to the value of harvested products quality and quantity, tend to be important components of silvicultural plans.

Good management plans include consideration of future conditions of the stand after any recommended harvests treatments, including future treatments (particularly in intermediate stand treatments), and plans for natural or artificial regeneration after final harvests.

The objectives of landowners and leaseholders influence plans for harvest and subsequent site treatment. In Britain, plans featuring "good forestry practice" must always consider the needs of other stakeholders such as nearby communities or rural residents living within or adjacent to woodland areas. Foresters consider tree felling and environmental legislation when developing plans. Plans instruct the sustainable harvesting and replacement of trees. They indicate whether road building or other forest engineering operations are required.

Agriculture and forest leaders are also trying to understand how the climate change legislation will affect what they do. The information gathered will provide the data that will determine the role of agriculture and forestry in a new climate change regulatory system.

Community Forestry

Community forestry is an evolving branch of forestry whereby the local community plays a significant role in forest management and land use decision making by themselves in the facilitating support of government as well as change agents. It involves the participation and collaboration of various stakeholders including community, government and non-government organisations (NGO's).

The level of involvement of each of these groups is dependent on the specific community forest project, the management system in use and the region. It gained prominence in the mid-1970s and examples of community forestry can now be seen in many countries including Nepal, Indonesia, Korea, Brazil, India and North America.

Community forestry is a branch of forestry that deals with the communal management of forests for generating income from timber and non-timber forest products as forms of goods while in other hand regulating ecosystem, downstream settlements benefits from watershed conservation, carbon sequestration and aesthetic values as in forms of services.

It has been considered one of the most promising options of combining forest conservation with rural development and community empowerment and poverty reduction objectives. Community forestry is defined by the Food and Agricultural Organization of the United Nations as "any situation that intimately involves local people in forestry activity". Community forestry exists when the local community in an area plays a significant role in land use decision-making and when the community is satisfied with its involvement and benefits from the management of the surrounding forest and its resources.

Community forestry is first implemented through the establishment of a legal and institutional framework including the revision of legal norms and regulations for forest management, the development of National

Forest Plans and the strengthening of decentralization processes to sub-national levels of government. The second principal line of action is the implementation of pilot projects to demonstrate the feasibility of the community forestry framework.

However, a study by the Overseas Development Institute shows that the technical, managerial and financial requirements stipulated by the framework are often incompatible with local realities and interests. A successful legal and institutional framework will incorporate the strengthening of existing institutions and enable the dissemination of locally appropriate practices as well as the local capacity for regulation and control.

History

Community forestry first came to prominence in the mid-1970s and has continued to evolve over the last few decades in a growing number of countries. The availability of forest resources are often greatly reduced for use by the local people due to increasing pressures to cultivate the land, reliance on the forest resources and are also affected by economic and political changes.

The evolution of community forestry in Nepal dates back to the late 1970s and was first instilled as an attempt to improve the management of forest resources and address environmental issues that were of great concern with the countries failing centralized forest policy. Over the past two decades, community forestry has been applied successfully in many developing countries, with its main goal being the alleviation of poverty amongst local forest communities and forest conservation. More recently, community forestry has been implemented in developing countries and it has been successful in its aims of sustainable forest management. Climate change adaptation plan of action and securing socio-economic benefits for local communities.

Challenges

A study conducted in the Brazilian Amazon determined that there are a number of challenges that must be faced when developing a sustainable management strategy for community forestry. The model is segregated into two phases: the 'development phase' during which several enabling factors (land ownership, organisational capacity, technical knowledge and capital) are needed to obtain a legal management permit and secondly the 'operational phase' where factors (clandestine loggers, access to markets, infrastructure and managerial skills) influence the successfulness of the management program.

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Graph Theory

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In mathematics and computer science, graph theory is the study of *graphs*, which are mathematical structures used to model pairwise relations between objects. A "graph" in this context is made up of "vertices" or "nodes" and lines called *edges* that connect them. A graph may be *undirected*, meaning that there is no distinction between the two vertices associated with each edge, or its edges may be *directed* from one vertex to another.

The graphs studied in graph theory should not be confused with the graphs of functions or other kinds of graphs.

Applications

Graphs can be used to model many types of relations and process dynamics in physical, biological, social and information systems. Many practical problems can be represented by graphs.

In computer science, graphs are used to represent networks of communication, data organization, computational devices, the flow of computation, etc. One practical example: The link structure of a website could be represented by a directed graph. The vertices are the web pages available at the website and a directed edge from page A to page B exists if and only if A contains a link to B. A similar approach can be taken to problems in travel, biology, computer chip design, and many other fields. The development of algorithms to handle graphs is therefore of major interest in computer science. There, the transformation of graphs is often formalized and represented by graph rewrite systems. They are either directly used or properties of the rewrite systems (e.g. confluence) are studied. Complementary to graph transformation systems focussing on rule-based in-memory manipulation of graphs are graph databases geared towards transaction-safe, persistent storing and querying of graph-structured data.

Graph-theoretic methods, in various forms, have proven particularly useful in linguistics, since natural language often lends itself well to discrete structure. Traditionally, syntax and compositional semantics follow tree-based structures, whose expressive power lies in the Principle of Compositionality, modelled in a hierarchical graph. More contemporary approaches such as Head-driven phrase structure grammar (HPSG) model syntactic constructions via the unification of typed feature structures, which are directed acyclic graphs.

Within lexical semantics, especially as applied to computers, modelling word meaning is easier when a given word is understood in terms of related words; semantic networks are therefore important in computational linguistics.

Still other methods in phonology (e.g. Optimality Theory, which uses lattice graphs) and morphology (e.g. finite-state morphology, using finite-state transducers) are common in the analysis of language as a graph. Indeed, the usefulness of this area of mathematics to linguistics has borne organizations such as TextGraphs, as well as various 'Net' projects, such as WordNet, VerbNet, and others.

Graph theory is also used to study molecules in chemistry and physics. In condensed matter physics, the three dimensional structure of complicated simulated atomic structures can be studied quantitatively by gathering statistics on graph-theoretic properties related to the topology of the atoms. For example, Franzblau's shortest-path (SP) rings. In chemistry a graph makes a natural model for a molecule, where vertices represent atoms and edges bonds. This approach is especially used in computer processing of molecular structures, ranging from chemical editors to database searching. In statistical physics, graphs can represent local connections between interacting parts of a system, as well as the dynamics of a physical process on such systems.

Graph theory is also widely used in sociology as a way, for example, to measure actors' prestige or to explore diffusion mechanisms, notably through the use of social network analysis software. Under the umbrella of Social Network graphs there are many different types of graphs: Starting with the Acquaintanceship and Friendship Graphs, these graphs are useful for representing whether no people know each other. Next there is the influence graph. This graph is used to model whether certain people can influence the behaviour of others. Finally there's a collaboration graph which models whether two people work together in a particular way. The measure of an actors' prestige mentioned above is an example of this, other popular examples include the Erdõs number and six degrees of separation

Likewise, graph theory is useful in biology and conservation efforts where a vertex can represent regions where certain species exist (or habitats) and the edges represent migration paths, or movement between the regions. This information is important when looking at breeding patterns or tracking the spread of disease, parasites or how changes to the movement can affect other species.

In mathematics, graphs are useful in geometry and certain parts of topology, e.g. Knot Theory. Algebraic graph theory has close links with group theory. A graph structure can be extended by assigning a weight to each edge of the graph. Graphs with weights, or weighted graphs, are used to represent structures in which pairwise connections have some numerical values. For example if a graph represents a road network, the weights could represent the length of each road.

A digraph with weighted edges in the context of graph theory is called a network. Network analysis have many practical applications, for example, to model and analyze traffic networks. Applications of network analysis split broadly into three categories:

1. First, analysis to determine structural properties of a network, such as the distribution of vertex degrees and the diameter of the graph. A vast number of graph measures exist, and the production of useful ones for various domains remains an active area of research.
2. Second, analysis to find a measurable quantity within the network, for example, for a transportation network, the level of vehicular flow within any portion of it.
3. Third, analysis of dynamical properties of networks.

History

The paper written by Leonhard Euler on the *Seven Bridges of Königsberg* and published in 1736 is regarded as the first paper in the history of graph theory. This paper, as well as the one written by Vandermonde on the *knight problem,* carried on with the *analysis situs* initiated by Leibniz. Euler's formula relating the number of edges, vertices, and faces of a convex polyhedron was studied and generalized by Cauchy and L'Huillier, and is at the origin of topology.

More than one century after Euler's paper on the bridges of Königsberg and while Listing introduced topology, Cayley was led by the study of particular analytical forms arising from differential calculus to study a particular class of graphs, the *trees*. This study had many implications in theoretical chemistry. The involved techniques mainly concerned the enumeration of graphs having particular properties. Enumerative graph theory then rose from the results of Cayley and the fundamental results published by Pólya between 1935 and 1937 and the generalization of these by De Bruijn in 1959. Cayley linked his results on trees with the contemporary studies of chemical composition. The fusion of the ideas coming from mathematics with those coming from chemistry is at the origin of a part of the standard terminology of graph theory.

In particular, the term "graph" was introduced by Sylvester in a paper published in 1878 in *Nature*, where he draws an analogy between "quantic invariants" and "co-variants" of algebra and molecular diagrams:

"[...] Every invariant and co-variant thus becomes expressible by a *graph* precisely identical with a Kekuléan diagram or chemicograph. [...] I give a rule for the geometrical multiplication of graphs, *i.e.* for constructing a *graph* to the product of in- or co-variants whose separate graphs are given. [...]" (italics as in the original).

The first textbook on graph theory was written by Dénes Kõnig, and published in 1936. A later textbook by Frank Harary, published in 1969, was enormously popular, and enabled mathematicians, chemists, electrical engineers and social scientists to talk to each other. Harary donated all of the royalties to fund the Pólya Prize.

One of the most famous and stimulating problems in graph theory is the four colour problem: "Is it true that any map drawn in the plane may have its regions coloured with four colours, in such a way that any two regions having a common border have different colours?" This problem was first posed by Francis Guthrie in 1852 and its first written record is in a letter of De Morgan addressed to Hamilton the same year. Many incorrect proofs have been proposed, including those by Cayley, Kempe, and others. The study and the generalization of this problem by Tait, Heawood, Ramsey and Hadwiger led to the study of the colourings of the graphs embedded on surfaces with arbitrary genus. Tait's reformulation generated a new class of problems, the *factorization problems*, particularly studied by Petersen and Kõnig. The works of Ramsey on colourations and more specially the results obtained by Turán in 1941 was at the origin of another branch of graph theory, *extremal graph theory*.

The four colour problem remained unsolved for more than a century. In 1969 Heinrich Heesch published a method for solving the problem using computers. A computer-aided proof produced in 1976 by Kenneth Appel and Wolfgang Haken makes fundamental use of the notion of "discharging" developed by Heesch. The proof involved checking the properties of 1,936 configurations by computer, and was not fully accepted at the time due to its complexity. A simpler proof considering only 633 configurations was given twenty years later by Robertson, Seymour, Sanders and Thomas.

The autonomous development of topology from 1860 and 1930 fertilized graph theory back through the works of Jordan, Kuratowski and Whitney. Another important factor of common development of graph theory and topology came from the use of the techniques of modern algebra. The first example of such a use comes from the work of the physicist Gustav Kirchhoff, who published in 1845 his Kirchhoff's circuit laws for calculating the voltage and current in electric circuits.

The introduction of probabilistic methods in graph theory, especially in the study of Erdõs and Rényi of the asymptotic probability of graph connectivity, gave rise to yet another branch, known as *random graph theory*, which has been a fruitful source of graph-theoretic results.

Graph Drawing

Graph drawing is an area of mathematics and computer science combining methods from geometric graph theory and information visualization to derive two-dimensional depictions of graphs arising from applications such as social network analysis, cartography, and bioinformatics.

A drawing of a graph or network diagram is a pictorial representation of the vertices and edges of a graph. This drawing should not be confused with the graph itself: very different layouts can correspond to the same graph. In the abstract, all that matters is which pairs vertices are connected by edges. In the concrete, however, the arrangement of these vertices and edges within a drawing affects its understandability, usability, fabrication cost, and aesthetics.

The problem gets worse, if the graph changes over time by adding and deleting edges (dynamic graph drawing) and the goal is to preserve the user's mental map.

Graphical Conventions

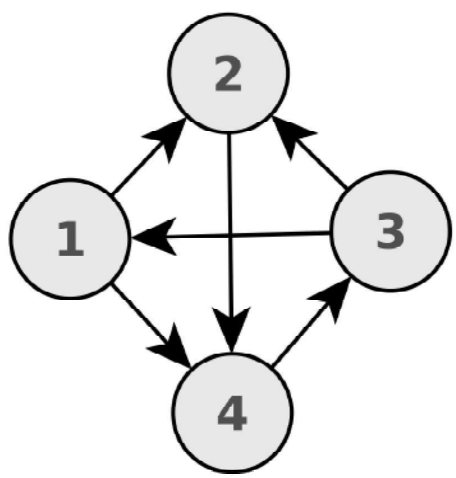

Figure: *Directed graph with arrowheads showing edge directions*

Graphs are frequently drawn as node-link diagrams in which the vertices are represented as disks or boxes and the edges are represented as line segments, polylines, or curves in the Euclidean plane.

In the case of directed graphs, arrowheads form a commonly used graphical convention to show their orientation; however, user studies have shown that other conventions such as tapering provide this information more effectively.

Alternative conventions to node-link diagrams include adjacency representations such as circle packings, in which vertices are represented by disjoint regions in the plane and edges are represented by adjacencies between regions; intersection representations in which vertices are represented by non-disjoint geometric objects and edges are represented by their intersections; visibility representations in which vertices are represented by regions in the plane and edges are represented by regions that have an unobstructed line of sight to each other; confluent drawings, in which edges are represented as smooth curves within mathematical train tracks; and visualizations of the adjacency matrix of the graph.

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An Overview of Dynamical System

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A dynamical system is a concept in mathematics where a fixed rule describes the time dependence of a point in a geometrical space. Examples include the mathematical models that describe the swinging of a clock pendulum, the flow of water in a pipe, and the number of fish each springtime in a lake. At any given time a dynamical system has a *state* given by a set of real numbers (a vector) that can be represented by a point in an appropriate *state space* (a geometrical manifold). Small changes in the state of the system create small changes in the numbers. The *evolution rule* of the dynamical system is a fixed rule that describes what future states follow from the current state. The rule is deterministic; in other words, for a given time interval only one future state follows from the current state.

The concept of a dynamical system has its origins in Newtonian mechanics. There, as in other natural sciences and engineering disciplines, the evolution rule of dynamical systems is an implicit relation that gives the state of the system for only a short time into the future. (The relation is either a differential equation, difference equation or other time scale.) To determine the state for all future times requires iterating the relation many times—each advancing time a small step. The iteration procedure is referred to as *solving the system* or *integrating the system*. Once the system can be solved, given an initial point it is possible to determine all its future positions, a collection of points known as a *trajectory* or *orbit*.

Before the advent of computers, finding an orbit required sophisticated mathematical techniques and could be accomplished only for a small class of dynamical systems. Numerical methods implemented on electronic computing machines have simplified the task of determining the orbits of a dynamical system. For simple dynamical systems, knowing the trajectory is often sufficient, but most dynamical systems are too complicated to be understood in terms of individual trajectories. The difficulties arise because:

- The systems studied may only be known approximately—the parameters of the system may not be known precisely or terms may be missing from the equations. The approximations used bring into question the validity or relevance of numerical solutions. To address these questions several notions of stability have been introduced in the study of dynamical systems, such as Lyapunov stability or structural stability. The stability of the dynamical system implies that there is a class of models or initial conditions for which the trajectories would be equivalent. The operation for comparing orbits to establish their equivalence changes with the different notions of stability.

- The type of trajectory may be more important than one particular trajectory. Some trajectories may be periodic, whereas others may wander through many different states of the system. Applications often require enumerating these classes or maintaining the system within one class. Classifying all possible trajectories has led to the qualitative study of dynamical systems, that is, properties that do not change under coordinate changes. Linear dynamical systems and systems that have two numbers describing a state are examples of dynamical systems where the possible classes of orbits are understood.
- The behaviour of trajectories as a function of a parameter may be what is needed for an application. As a parameter is varied, the dynamical systems may have bifurcation points where the qualitative behaviour of the dynamical system changes. For example, it may go from having only periodic motions to apparently erratic behaviour, as in the transition to turbulence of a fluid.
- The trajectories of the system may appear erratic, as if random. In these cases it may be necessary to compute averages using one very long trajectory or many different trajectories. The averages are well defined for ergodic systems and a more detailed understanding has been worked out for hyperbolic systems. Understanding the probabilistic aspects of dynamical systems has helped establish the foundations of statistical mechanics and of chaos.

It was in the work of Poincaré that these dynamical systems themes developed.

Examples

The evolution function Φ^t is often the solution of a *differential equation of motion*

$$\dot{x} = v(x).$$

The equation gives the time derivative, represented by the dot, of a trajectory $x(t)$ on the phase space starting at some point x_0. The *vector field* $v(x)$ is a smooth function that at every point of the phase space M provides the velocity vector of the dynamical system at that point. (These vectors are not vectors in the phase space M, but in the tangent space T_xM of the point x.) Given a smooth Φ^t, an autonomous vector field can be derived from it.

There is no need for higher order derivatives in the equation, nor for time dependence in $v(x)$ because these can be eliminated by considering systems of higher dimensions. Other types of differential equations can be used to define the evolution rule:

$$G(x, \dot{x}) = 0$$

is an example of an equation that arises from the modelling of mechanical systems with complicated constraints.

The differential equations determining the evolution function Φ^t are often ordinary differential equations: in this case the phase space M is a finite dimensional manifold. Many of the concepts in dynamical systems can be extended to infinite-dimensional manifolds—those that are locally Banach spaces—in which case the differential equations are partial differential equations. In the late 20th century the dynamical system perspective to partial differential equations started gaining popularity.

Linear Dynamical Systems

Linear dynamical systems can be solved in terms of simple functions and the behaviour of all orbits classified. In a linear system the phase space is the N-dimensional Euclidean space, so any point in phase space can be represented by a vector with N numbers. The analysis of linear systems is possible because they satisfy a superposition principle: if $u(t)$ and $w(t)$ satisfy the differential equation for the vector field (but not necessarily the initial condition), then so will $u(t) + w(t)$.

Flows

For a flow, the vector field $\Phi(x)$ is an affine function of the position in the phase space, that is,

$$\dot{x} = \phi(x) = Ax + b,$$

with A a matrix, b a vector of numbers and x the position vector. The solution to this system can be found by using the superposition principle (linearity). The case $b \neq 0$ with $A = 0$ is just a straight line in the direction of b:

$$\Phi^t(x_1) = x_1 + bt.$$

When b is zero and $A \neq 0$ the origin is an equilibrium (or singular) point of the flow, that is, if $x_0 = 0$, then the orbit remains there. For other initial conditions, the equation of motion is given by the exponential of a matrix: for an initial point x_0,

$$\Phi^t(x_0) = e^{tA} x_0.$$

When $b = 0$, the eigenvalues of A determine the structure of the phase space. From the eigenvalues and the eigenvectors of A it is possible to determine if an initial point will converge or diverge to the equilibrium point at the origin.

The distance between two different initial conditions in the case $A \neq 0$ will change exponentially in most cases, either converging exponentially fast towards a point, or diverging exponentially fast. Linear systems display sensitive dependence on initial conditions in the case of divergence. For nonlinear systems this is one of the (necessary but not sufficient) conditions for chaotic behaviour.

Maps

A discrete-time, affine dynamical system has the form

$$x_{n+1} = Ax_n + b,$$

with A a matrix and b a vector. As in the continuous case, the change of coordinates $x \to x + (1-A)^{-1} b$ removes the term b from the equation. In the new coordinate system, the origin is a fixed point of the map and the solutions are of the linear system $A^n x_0$. The solutions for the map are no longer curves, but points that hop in the phase space. The orbits are organized in curves, or fibres, which are collections of points that map into themselves under the action of the map.

As in the continuous case, the eigenvalues and eigenvectors of A determine the structure of phase space. For example, if u_1 is an eigenvector of A, with a real eigenvalue smaller than one, then the straight lines given by the points along αu_1, with $\alpha \in \mathrm{R}$, is an invariant curve of the map. Points in this straight line run into the fixed point.

Local Dynamics

The qualitative properties of dynamical systems do not change under a smooth change of coordinates (this is sometimes taken as a definition of qualitative): a singular point of the vector field (a point where v(x) = 0) will remain a singular point under smooth transformations; a periodic orbit is a loop in phase space and smooth deformations of the phase space cannot alter it being a loop. It is in the neighbourhood of singular points and periodic orbits that the structure of a phase space of a dynamical system can be well understood. In the qualitative study of dynamical systems, the approach is to show that there is a change of coordinates (usually unspecified, but computable) that makes the dynamical system as simple as possible.

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Mechanisms of Reproductive Isolation

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The mechanisms of reproductive isolation or hybridization barriers are a collection of mechanisms, behaviours and physiological processes that prevent the members of two different species that cross or mate from producing offspring, or which ensure that any offspring that may be produced are sterile. These barriers maintain the integrity of a species over time, reducing or directly impeding gene flow between individuals of different species, allowing the conservation of each species' characteristics. The mechanisms of reproductive isolation have been classified in a number of ways. Zoologist Ernst Mayr classified the mechanisms of reproductive isolation in two broad categories: those that act before fertilization (or before mating in the case of animals, which are called pre-copulatory) and those that act after. These have also been termed pre-zygotic and post-zygotic mechanisms. The different mechanisms of reproductive isolation are genetically controlled and it has been demonstrated experimentally that they can evolve in species whose geographic distribution overlaps (sympatric speciation) or as the result of adaptive divergence that accompanies allopatric speciation.

Pre-zygotic isolation mechanisms are the most economic in terms of the biological efficiency of a population, as resources are not wasted on the production of a descendent that is weak, non-viable or sterile.

Temporal or Habitat Isolation

Any of the factors that prevent potentially fertile individuals from meeting will reproductively isolate the members of distinct species. The types of barriers that can cause this isolation include: different habitats, physical barriers, and a difference in the time of sexual maturity or flowering. When factors change, especially physical barriers, often, species will branch off.

An example of the ecological or habitat differences that impede the meeting of potential pairs occurs in two fish species of the family *Gasterosteidae* (sticklebacks). One species lives all year round in fresh water, mainly in small streams. The other species lives in the sea during winter, but in spring and summer individuals migrate to river estuaries to reproduce. The members of the two populations are reproductively isolated due to their adaptations to distinct salt concentrations. An example of reproductive isolation due to differences in the mating season are found in the toad species *Bufo americanus* and *Bufo fowleri* . The members of these species can be successfully crossed in the laboratory producing healthy, fertile hybrids. However, mating does not occur in the wild even though the geographical distribution of the two species overlaps. The reason for the absence of inter-species mating is that *B. americanus* mates in early summer and *B. fowleri* in late summer. Certain plant species, such as *Tradescantia canaliculata* and *T. subaspera*, are sympatric throughout their geographic distribution yet they are reproductively isolated as they flower at different times of the year. In addition, one species grows in sunny areas and the other in deeply shaded areas.

Sexual Isolation by Behaviour or Conduct

The different mating rituals of animal species creates extremely powerful reproductive barriers, termed sexual or behaviour isolation, that isolate apparently similar species in the majority of the groups of the animal kingdom. In dioecious species, males and females have to search for a partner, be in proximity to each other, carry out the complex mating rituals and finally copulate or release their gametes into the environment in order to breed.

Pheromones play an important role in the sexual isolation of insect species. These compounds serve to identify individuals of the same species and of the same or different sex. Evaporated molecules of volatile pheromones can serve as a wide-reaching chemical signal. In other cases, pheromones may be detected only at a short distance or by contact.

In species of the *melanogaster* group of *Drosophila*, the pheromones of the females are mixtures of different compounds, there is a clear dimorphism in the type and/or quantity of compounds present for each sex. In addition, there are differences in the quantity and quality of constituent compounds between related species, it is assumed that the pheromones serve to distinguish between individuals of each species. An example of the role of pheromones in sexual isolation is found in 'corn borers' in the genus *Ostrinia*. There are two twin species in Europe that occasionally cross. The females of both species produce pheromones that contain a volatile compound which has two isomers, E and Z; 99% of the compound produced by the females of one species is in the E isomer form, while the females of the other produce 99% isomer Z. The production of the compound is controlled by just one locus and the interspecific hybrid produces an equal mix of the two isomers. The males, for their part, almost exclusively detect the isomer emitted by the females of their species, such that the hybridization although possible is scarce. The perception of the males is controlled by one gene, distinct from the one for the production of isomers, the heterozygous males show a moderate response to the odour of either type. In this case, just 2 'loci' produce the effect of ethological isolation between species that are genetically very similar.

Sexual isolation between two species can be asymmetrical. This can happen when the mating that produces descendants only allows one of the two species to function as the female progenitor and the other as the male, while the reciprocal cross does not occur. For instance, half of the wolves tested in the Great Lakes area of America show mitochondrial DNA sequences of coyotes, while mitochondrial DNA from wolves is never found in coyote populations. This probably reflects an asymmetry in inter-species mating due to the difference in size of the two species as male wolves take advantage of their greater size in order to mate with female coyotes, while female wolves and male coyotes do not mate

Mechanical Isolation

Mating pairs may not be able to couple successfully if their genitals are not compatible. The relationship between the reproductive isolation of species and the form of their genital organs was signaled for the first time in 1844 by the French entomologist Léon Dufour. Insects' rigid carapaces act in a manner analogous to a lock and key, as they will only allow mating between individuals with complementary structures, that is, males and females of the same species (termed *co-specifics*).

Evolution has led to the development of genital organs with increasingly complex and divergent characteristics, which will cause mechanical isolation between species. Certain characteristics of the genital organs will often have converted them into mechanisms of isolation. However, numerous studies show that organs that are anatomically very different can be functionally compatible, indicating that other factors also determine the form of these complicated structures.

Mechanical isolation also occurs in plants and this is related to the adaptation and coevolution of each species in the attraction of a certain type of pollinator (where pollination is zoophilic) through a collection of morphophysiological characteristics of the flowers (called floral syndrome), in such a way that the transport of pollen to other species does not occur.

Gametic Isolation

The synchronous spawning of many species of coral in marine reefs means that inter-species hybridization can take place as the gametes of hundreds of individuals of tens of species are liberated into the same water at the same time. Approximately a third of all the possible crosses between species are compatible, in the sense that the gametes will fuse and lead to individual hybrids. This hybridization apparently plays a fundamental role in the evolution of coral species. However, the other two-thirds of possible crosses are incompatible. It has been observed that in sea urchins of the genus *Strongylocentrotus* the concentration of spermatocytes that allow 100% fertilization of the ovules of the same species is only able to fertilize 1.5% of the ovules of other species. This inability to produce hybrid offspring, despite the fact that the gametes are found at the same time and in the same place, is due to a phenomenon known as *gamete incompatibility*, which is often found between marine invertebrates, and whose physiological causes are not fully understood.

A number of mechanisms which act after fertilisation preventing successful inter-population crossing.

Reproductive isolation between species appears, in certain cases, a long time after fertilization and the formation of the zygote, as happens - for example - in the twin species *Drosophila pavani* and *D. gaucha*. The hybrids between both species are not sterile, in the sense that they produce viable gametes, ovules and spermatozoa. However, they cannot produce offspring as the sperm of the hybrid male do not survive in the semen receptors of the females, be they hybrids or from the parent lines. In the same way, the sperm of the males of the two parent species do not survive in the reproductive tract of the hybrid female.

The hybrids of two populations with differing numbers of chromosomes can experience a certain loss of fertility, and therefore a poor adaptation, because of irregular meiosis.

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Route Problems

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Hamiltonian Path Problem

In the mathematical field of graph theory the Hamiltonian path problem and the Hamiltonian cycle problem are problems of determining whether a Hamiltonian path or a Hamiltonian cycle exists in a given graph (whether directed or undirected). Both problems are NP-complete.

Relation Between Problems

There is a simple relation between the problems of finding a Hamiltonian path and a Hamiltonian cycle. In one direction, the Hamiltonian path problem for graph G is equivalent to the Hamiltonian cycle problem in a graph H obtained from G by adding a new vertex and connecting it to all vertices of G. Thus, finding a Hamiltonian path cannot be significantly slower (in the worst case, as a function of the number of vertices) than finding a Hamiltonian cycle. In the other direction, a graph G has a Hamiltonian cycle using edge uv if and only if the graph H obtained from G by replacing the edge by a pair of vertices of degree 1, one connected to u and one connected to v, has a Hamiltonian path.

Therefore, by trying this replacement for all edges incident to some chosen vertex of G, the Hamiltonian cycle problem can be solved by at most n Hamiltonian path computations, where n is the number of vertices in the graph. The Hamiltonian cycle problem is also a special case of the travelling salesman problem, obtained by setting the distance between two cities to one if they are adjacent and two otherwise, and verifying that the total distance travelled is equal to n (if so, the route is a Hamiltonian circuit; if there is no Hamiltonian circuit then the shortest route will be longer).

Algorithms

There are $n!$ different sequences of vertices that might be Hamiltonian paths in a given n-vertex graph (and are, in a complete graph), so a brute force search algorithm that tests all possible sequences would be very slow. There are several faster approaches. A search procedure by Frank Rubin divides the edges of the graph into three classes: those that must be in the path, those that cannot be in the path, and undecided. As the search proceeds, a set of decision rules classifies the undecided edges, and determines whether to halt or continue the search.

The algorithm divides the graph into components that can be solved separately. Also, a dynamic programming algorithm of Bellman, Held, and Karp can be used to solve the problem in time $O(n^2 2^n)$.

In this method, one determines, for each set S of vertices and each vertex v in S, whether there is a path that covers exactly the vertices in S and ends at v. For each choice of S and v, a path exists for (S,v) if and only if v has a neighbour w such that a path exists for $(S - v,w)$, which can be looked up from already-computed information in the dynamic programme. Andreas Björklund provided an alternative approach using the inclusion–exclusion principle to reduce the problem of counting the number of Hamiltonian cycles to a simpler counting problem, of counting cycle covers, which can be solved by computing certain matrix determinants.

Using this method, he showed how to solve the Hamiltonian cycle problem in arbitrary n-vertex graphs by a Monte Carlo algorithm in time $O(1.657^n)$; for bipartite graphs this algorithm can be further improved to time $O(1.414^n)$. For graphs of maximum degree three, a careful backtracking search can find a Hamiltonian cycle (if one exists) in time $O(1.251^n)$.

Because of the difficulty of solving the Hamiltonian path and cycle problems on conventional computers, they have also been studied in unconventional models of computing. For instance, Leonard Adleman showed that the Hamiltonian path problem may be solved using a DNA computer. Exploiting the parallelism inherent in chemical reactions, the problem may be solved using a number of chemical reaction steps linear in the number of vertices of the graph; however, it requires a factorial number of distinct types of DNA molecule to participate in the reaction.

Complexity

The problem of finding a Hamiltonian cycle or path is in FNP; the analogous decision problem is to test whether a Hamiltonian cycle or path exists. The directed and undirected Hamiltonian cycle problems were two of Karp's 21 NP-complete problems. They remain NP-complete even for undirected planar graphs of maximum degree three, for directed planar graphs with indegree and outdegree at most two, for bridgeless undirected planar 3-regularbipartite graphs, and for 3-connected 3-regular bipartite graphs. However, putting all of these conditions together, it remains open whether 3-connected 3-regular bipartite planar graphs must always contain a Hamiltonian cycle, in which case the problem restricted to those graphs could not be NP-complete.

In graphs in which all vertices have odd degree, an argument related to the handshaking lemma shows that the number of Hamiltonian cycles through any fixed edge is always even, so if one Hamiltonian cycle is given, then a second one must also exist.

However, finding this second cycle does not seem to be an easy computational task. Papadimitriou defined the complexity class PPA to encapsulate problems such as this one.

Minimum Spanning Tree

Given a connected, undirected graph, a spanning tree of that graph is a subgraph that is a tree and connects all the vertices together. A single graph can have many different spanning trees. We can also assign a *weight* to each edge, which is a number representing how unfavourable it is, and use this to assign a weight to a spanning tree by computing the sum of the weights of the edges in that spanning tree. A minimum spanning tree (MST) or minimum weight spanning tree is then a spanning tree with weight less than or equal to the weight of every other spanning tree.

More generally, any undirected graph (not necessarily connected) has a minimum spanning forest, which is a union of minimum spanning trees for its connected components.

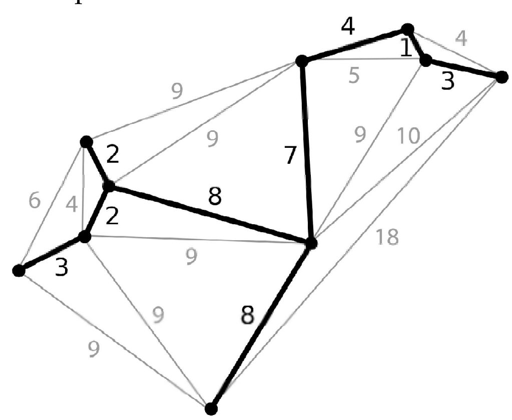

Figure: *The only minimum spanning tree of a planar graph. Each edge is labelled with its weight, which here is roughly proportional to its length.*

One example would be a telecommunications company laying cable to a new neighbourhood. If it is constrained to bury the cable only along certain paths, then there would be a graph representing which points are connected by those paths. Some of those paths might be more expensive, because they are longer, or require the cable to be buried deeper; these paths would be represented by edges with larger weights.

A *spanning tree* for that graph would be a subset of those paths that has no cycles but still connects to every house. There might be several spanning trees possible.

A *minimum spanning tree* would be one with the lowest total cost.

Properties

Possible multiplicity:

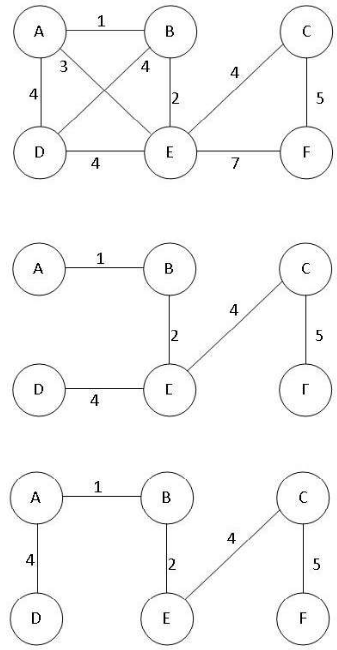

Figure: *This figure shows there may be more than one minimum spanning tree in a graph. In the figure, the two trees below the graph are two possibilities of minimum spanning tree of the given graph.*

There may be several minimum spanning trees of the same weight having a minimum number of edges; in particular, if all the edge weights of a given graph are the same, then every spanning tree of that graph is minimum. If there are *n* vertices in the graph, then each tree has *n*-1 edges.

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A New Species of MicroveliaWestwood,1834 from India

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Abstract: *Microvelia*Westwood,1834 is an important genus of family *Veliidae.*1843.It is hereto represented by seven species viz. *M.repentina* Distant,1904, *M. singalensis* Kirkaldy,1903,M. albomaculata Distant,1909, *M.kumaonensis* Distant,1909, *M.diluta* Distant 1909 and *M. annandelei* Distant, 1909. One new species of the genus *Microvelia(s.str.)andersoni* sp.nov. is described from Mathura, India

The *Vellidae* are best known of all aquatic hemiptera of the world and are extremely common in Indian waters. The members of the family are easily differentiated from the related family *Gerridae* based on hind leg not surpassing beyond the tip of abdomen and the presence of median longitudinal groove on vertex.*Microvelia* Westwood,1834 belongs to the subfamily Microveliinae China & Usinger, 1949 of the family *Veliidae.* The genus *Microvelia* Westwood, 1834 are easily recognized from the members of other genera due to subangular and posteriorly produced pronotum.

Genus-*Microvelia* Westwood, 1834

Type species: *Microvelia pulchella* Westwood, 1834

Type: British Museum,London

Disribution: The Oriental realm (Burma, Ceylon, India, Java, Malaya, Philippines), Australian realm, Southern Eastern United State into Central Mexico. Japan(Honshu and Kyushu)

Microvelia andersoni sp.nov.

Description: Size: Male, Apterous,2.12-2.22 mm length, width of head across eyes 0.43-0.44mm; Female, Apterous 2.31-2.42mm in length, width across metathorax 0.35-1.02mm.

Colour: Apterous male brown with silvery hair patches. Head blackish .Eyes dark red.Antennae dark brown in colour. Pronotum black with bright yellowish spots and dark punctures. Pro and mesosternum dark brownish. Abdominal tergites and latero tergites pale .Apterous female similar with male except abdomen with few black dark spots.

Structural Charcteristics:

Head: Head across eyes about two third as long as wide (24:41 in Apterous male and 21:32 in Apterous female). Eyes small, antennae relatively long and slender, about half the total length. Relative length of antennal segments of Apterous male Ist:IInd:IIIrd:IVth : :23:15:21:28 and of female Ist:IInd:IIIrd:IVth :: 20: 15:20:28. Antenniferous tubercle poorly developed. Basal margin of clypeus distinct.Mandibullar and Maxillary plates well separated with each other. Rostrum slender, reaching middle of mesosternum, third segment a little more than twice the length of last segment (5:2 in male and 5:1 in female)

Thorax: Pronotum long(21:66 in Apterous male and 25:55 in Apterous female). Pronotum broad round. Intersegmental suture between Mesonotum and meganotum well demarcated with median longitudinal sulcus. Legs relatively long and slender.

Relative Length of Leg Segments:

Apterous Male (2.12-2.22mm)

Femur	*tibia*	*tarsus*	*First tarsal segment*	*Second tarsal segment*
Fore leg	50	40	20	- -
Mid leg	60	50	30	11 18
Hind leg	72	73	34	11 21

Apterous Female (2.31-2.42 mm)

Femur	*tibia*	*tarsus*	*First tarsal segment*	*Second tarsal segment*
Fore leg	47	37	22	- -
Mid leg	57	57	31	1019
Hind leg	72	76	32	1120

Abdomen: Abdominal tergites together a little more than three times as long as pronotum (119: 34 in male and 126:30 in female) 7th tergite longer than first two segment (35:29 in male and 40:30 in female). Connexivum slightly raised above. Seventh sternum with apical margin slightly concave with deep pit in the middle. Oval tubercle with many denticles situiated between the raised margins.

Male genitalia: Eighth segment a little longer than seventh tergite(41:37), ventral apical margin concave with a long process, slightly curved with numerous denticles . Ninth segment relatively short, slightly assymetrical. Claspers nearly identical,small,subconical with numerous small hairs on the antero-dorsally directed surface. Phallus with a very complicated set of ill-defined sclerite.

Female genitalia: Genital segment completely withdrawn into cylindrical 7th abdominal segment and bent downward , covering velvifers. Eight segment with first valvula long,narrow, rounded apex, second valvula with lateral margin broadly sclerotized, narrow, long extending beyond apical margin of intervalvular membrane. Apical process of intervalvular membrane pointed

Material examined: *Holotype* Apterous one female, *allotype* Apterous one male on pins. *Paratype* Apterous 3 males,3 females in spirit.Mathura 21.VIII.2012 (Anshul).River Yamuna 2 males and 2 apterous females collected from Compu ghat area Mathura12.X.2012 (Ramhans)

Distribution: INDIA: UTTAR PRADESH

Remark: This species is closely related to *Microvelia mjobergi* Hale (1925) but somewhat larger than *Microvelia andersoni* sp. nov. The vertical process of 8TH segment is broad basally with a slender and pointed apex not evenly tapering as in *Microvelia andersoni*sp.nov.

The species is named after Dr.N.M. Anderson for his outstanding contribution in the study of Hemiptera

Acknowledgement: The author is greatful to Dr.Ajay Tyagi, Principal K.R.(P.G.) College, Mathura for providing necessary library and laboratory facilities. Thanks are due to Dr. Y.C. Gupta ,Former HOD and Director B.S.A. College, Mathura for valuable suggestions.

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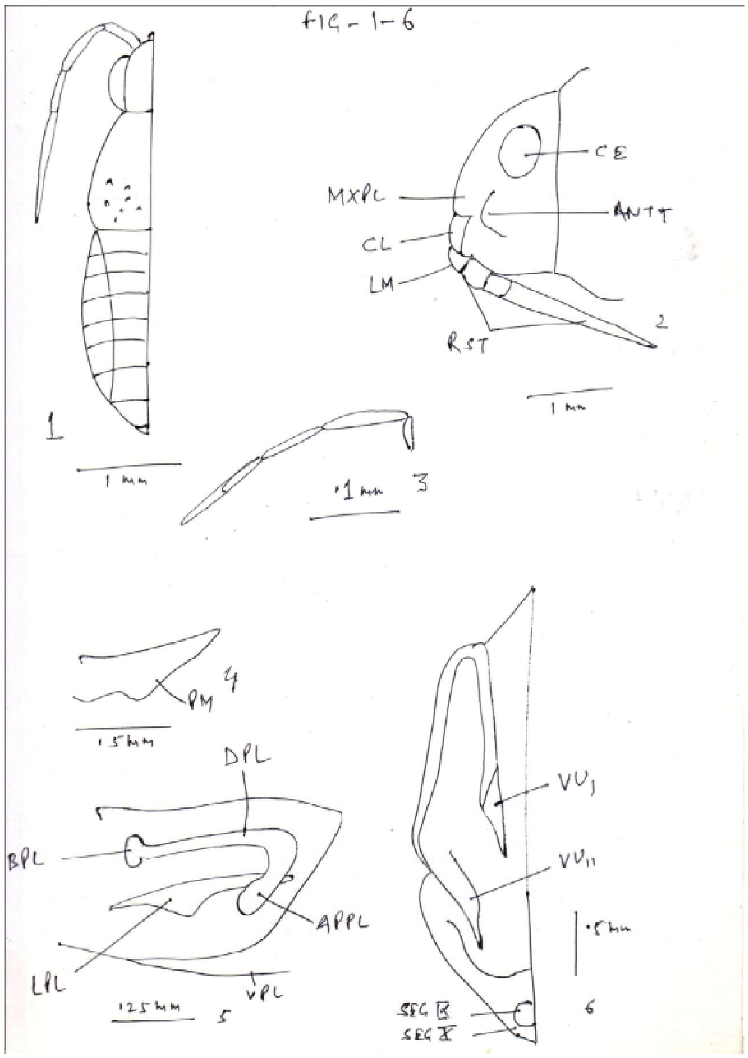
FIG - 1-6
1
1 mm
CE
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3
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LPL
VPL
·125 mm
5
VU1
VU11
·5 mm
SEG IX
SEG X
6

Bioremediation of Heavy Metals Using the Aquatic Plant Salvinia

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Abstract

Salvinia guettardii is classified as a Weed of National Significance and is considered one of the most problematic weeds due to its invasive nature, potential for widespread proliferation, and significant economic and environmental consequences. This aquatic weed can obstruct waterways, as it floats on still or slow-moving water and can rapidly expand to cover the entire surface with a dense layer of vegetation. This growth blocks sunlight from reaching submerged plants and disrupts oxygen exchange, rendering the water inhospitable for fish and other aquatic organisms. The presence of heavy metal toxicity and the risk of bioaccumulation in the food chain pose serious environmental and health challenges in contemporary society. A promising, environmentally friendly, and cost-effective solution is plant-based bioremediation, also known as phytoremediation. Aquatic ferns, in particular, demonstrate exceptional capabilities in removing various contaminants, including heavy metals, organic compounds, and radionuclides from the environment. Among the diverse aquatic macrophytes, Salvinia, a free-floating fern, stands out due to its high productivity and ability to thrive in a wide range of temperatures. This study focused on the bioremediation potential of three Salvinia species—S. natans, S. molesta, and S. auriculata—regarding heavy metal contaminants. The findings revealed that varying concentrations of heavy metals led to a significant decrease in the fresh weight and an increase in the dry weight of all three species. All three species of Salvinia exhibited a reduction in fresh weight as the concentrations of Cd, Cu, Cr, Hg, Pb, Ni, and Zn increased. After a treatment period of 10 days, these species demonstrated a significant accumulation of heavy metals within their tissues. The bioaccumulation of heavy metals was found to rise in correlation with the increasing concentrations of these metals. Salvinia molesta displayed the highest levels of accumulation for Hg, Ni, and Pb, with concentrations reaching 18575 ppm, 18875 ppm, and 18275 ppm, respectively. It can be concluded that Salvinia natans, S. molesta, and S. auriculata are highly effective in accumulating substantial amounts of heavy metals. The findings from this experimental study confirm that Salvinia natans, S. molesta, and S. auriculata possess a natural ability to accumulate significant quantities of heavy metals.

Keywords: Heavy metals, phytoremediation, bioaccumulation, hyperaccumulators, Salvinia.

Introduction

Salvinia is a type of free-floating water fern that creates thick mats on the surface of water bodies. It is characterized by numerous branched horizontal stems, measuring 1–2 mm in diameter, which float just beneath the water's surface. Each node produces a pair of green, oval-shaped hairy fronds that float. Additionally, a brown frond, composed of many hairy filaments, emerges from each node and extends into the water, resembling and functioning like a root. As Salvinia matures and becomes denser, its appearance changes. In sparse populations, the primary invasive plants feature a few small floating fronds (10–15 mm wide) that lie flat on the water. Conversely, in dense infestations that form mats, the numerous floating fronds can expand up to 60 mm wide, folding and overlapping in a tightly packed, concertina-like arrangement. The surfaces of these floating fronds are adorned with numerous distinctive egg-beater-shaped hairs that help repel water and enhance buoyancy.

Salvinia has the ability to increase its density through growth, with stems capable of reaching lengths of up to 300 mm, or through vegetative reproduction. This plant proliferates at an alarming rate, with infestations potentially doubling in size every two to three days. Areas that are not yet infested can rapidly become entirely overrun by salvinia, even with the introduction of only small quantities into a waterway. Reproduction takes place when mature plants generate buds at the stem nodes, which are the junctions between stem sections, leading to the formation of new daughter plants. Additionally, if a segment of the stem containing a node detaches from the main plant, it can develop into a new individual. A single pair of fronds has the potential to initiate an entirely new infestation. Salvinia can easily spread downstream during flooding events, but it can also invade new catchments through human activities. It has been deliberately disseminated globally as a decorative plant for ponds and aquariums, and it has frequently either escaped or been introduced into various waterways. Additionally, it can be inadvertently introduced into new water bodies when it clings to boats and other aquatic gear.

Weed mats obstruct swimming and render fishing impossible. The displacement of native aquatic plants, birds, and animals can tarnish the natural beauty of open water bodies, leading to further degradation. During flood events, accumulations of weed material gather at fences and bridges, which subsequently trap additional floating debris. The resulting weight can lead to the collapse of these structures. Additionally, the roots of the weeds restrict water flow to irrigation systems, resulting in increased pumping times and costs. Consequently, Salvinia is classified as a Weed of National Significance.

Salvinia comprises thirteen distinct species. Each species of Salvinia is a free-floating aquatic fern characterized by small, buoyant green leaves that are arranged in pairs along a shared stem. The surface of each leaf is adorned with long, rigid, water-repellent hairs. As the plant reaches maturity, the leaves thicken and bend at the mid-rib. Species such as Salvinia natans, S. molesta, and S. auriculata are present in India. Salvinia natans is particularly abundant in Dal Lake located in Kashmir (Srinagar), while Salvinia molesta is found in the state of Kerala. This genus thrives in warmer climates and is sensitive to frost, resulting in minimal growth during the winter months. However, with the rise in summer temperatures, Salvinia experiences a significant increase in vegetative growth. Under ideal conditions, Salvinia can double its volume every two to three days.[1]

Water pollution represents a significant global challenge and is a leading contributor to mortality and illness worldwide.[2] As noted by M.A. Farooqui, a scientist with the Central Ground Water Board (CGWB) in 2011, the improper disposal of solid waste directly contaminates groundwater sources. The most potent water pollutants include insecticides, waste from livestock operations, volatile organic compounds, byproducts from food processing, and various chemical wastes. Among these, heavy metals pose the greatest risk to health, with Cadmium (Cd), Chromium (Cr), Copper (Cu), Mercury (Hg), Lead (Pb), Nickel (Ni), and Zinc (Zn) being the most prevalent contaminants (Lasat, 2002). Industrial wastewater is a significant contributor to heavy metal contamination. Chromium and copper are the primary elements involved, resulting in the degradation of numerous aquatic ecosystems. Chromium in its hexavalent form (Cr (VI)) is recognized as the most hazardous variant, typically found in conjunction with oxygen as chromate (CrO4 2-) or dichromate (Cr2O7 2-) oxyanions. In contrast, trivalent chromium (Cr (III)) is less mobile and toxic, primarily existing in association with organic matter in soil and water environments (Bequer et al., 2003). In India, approximately 2,000 to 32,000 tons of elemental chromium are released into the environment each year due to the activities of tanning industries. Copper (Cu) levels exceeding 20 micrograms per gram (ìg/g) can be harmful, as noted by Heike Bradl (2005) and Wright and Welbourn (2002).[3] The toxicity of heavy metals and the risk of their accumulation in the food chain pose significant environmental and public health challenges in contemporary society. Traditional methods for the removal of metal ions are often prohibitively expensive. An effective, environmentally friendly, and cost-efficient alternative is plant-based bioremediation, also known as phytoremediation, which was initially introduced by R.L. Chaney in 1983. Phytoremediation describes the inherent capability of specific plants to bioaccumulate, decompose, or neutralize contaminants present in soil, water, or air through various natural biological, chemical, or physical processes. Certain aquatic plants are categorized as 'hyperaccumulators.' Notably, aquatic ferns demonstrate exceptional capacity to eliminate a range of pollutants, including heavy metals, organic compounds, and radionuclides from the environment. Among these species, Salvinia, a free-floating aquatic fern, stands out due to its high productivity and resilience to a broad spectrum of temperatures. Pistia stratiotes L. demonstrates significant potential for bioaccumulation and serves as a bioindicator for a range of heavy metals [4].

The significance of aquatic plants in phytoremediation technology is well recognized. Notably, aquatic ferns demonstrate remarkable capabilities in eliminating a range of pollutants, including heavy metals, organic substances, and radionuclides from the environment[5].

Salvinia occupies a unique role among various species due to its numerous advantages, such as high productivity and the ability to withstand a broad range of temperatures (Olguín et al., 2002). Several Salvinia species, including S. herzogii, S. minima, S. natans, and S. rotundifolia, demonstrate significant potential for the removal of various contaminants, including heavy metals, from wastewater.[6] The capacity of Salvinia to eliminate heavy metals has been the subject of extensive research. The removal and compartmentalization of heavy metals in Salvinia largely depend on the availability of specific nutrients and chelating agents, with environmental factors playing a secondary role. The mechanism of metal uptake varies according to the species of the plant and the type of metal involved. In Salvinia, metal uptake occurs through both biological and physical mechanisms. The physical uptake of metals such as chromium (Cr) and lead (Pb) is rapid and involves processes like adsorption, ionic exchange, and chelation. In contrast, biological uptake, which includes intracellular transport through the plasmalemma into the cells, is slower but facilitates the subsequent movement of metals like cadmium (Cd) from the roots to the leaves.[7] The highest absorption rate is observed within the initial hours; however, the sorption capacity is constrained by the availability of adsorption sites. Research utilizing scanning electron microscopy microanalysis indicates that heavy metals are directly absorbed through leaves, as they are in direct contact with the solution. This direct sorption is suggested to be the primary factor contributing to the increased metal concentration in the aerial parts of the plants. It has been theorized that the uptake of heavy metals is facilitated by secondary transport proteins, such as channel proteins or H+ coupled carrier proteins, where the negative membrane potential within the plasma membrane promotes the uptake of cations via secondary transporters. The presence of free carboxylic groups on the cell surface serves as binding sites for metals (Olguin et al., 2005). The significant metal removal capacity of Salvinia biomass is attributed to its large specific surface area (264 m^2 g"1), which is abundant in carbohydrates (48.50%) and carboxyl groups (0.95 mmol g"1). Proteins serve as significant ligand atoms and are crucial for the absorption of metals. The kinetics associated with metal removal demonstrate a first-order rate, and the equilibrium data align well with both Langmuir and Freundlich isotherms. Among the various species of Salvinia, S. minima is recognized as a hyperaccumulator of lead and cadmium due to its elevated bioconcentration factor (BCF), which can range from 2000 to 2600 in batch systems and from 4134 to 17170 in continuous systems. The non-living biomass of Salvinia also shows a similarly high capacity for heavy metal removal. The increased levels of lipids and carbohydrates on the plant's surface function as cationic weak exchanger groups, facilitating metal sorption through ion exchange reactions. The sorption of heavy metals by dry biomass adheres to the Langmuir isotherm as well.[8]

Materials and Methods

Uniform specimens of three Salvinia species, namely Salvinia natans, S. molesta, and S. auriculata, were gathered from various lakes located in Srinagar and Kerala. Specifically, S. natans was sourced from Srinagar, while S. molesta and S. auriculata were obtained from Kerala, ensuring consistency in size, mass, and root length among the samples.

The collected plants were thoroughly washed multiple times with tap water, followed by a final rinse with deionized distilled water to eliminate impurities such as periphyton, dust, and sediment particles. The cleaned material was then stored in polythene bags. Concurrently, water samples from the vicinity of the plants were randomly collected and transported to the laboratory, where the temperature of the water at the time of sampling was noted. Each individual plant sample underwent an additional wash with distilled deionized water in the laboratory.

To prepare the molar stock solutions of various heavy metal salts, specific amounts were dissolved in 1000 ml of distilled water for each compound: 138.32g of Cadmium chloride, 294.185g of potassium dichromate, 159.609g of Copper sulphate, 271.52g of Mercuric chloride, 278.1g of Lead chloride, 129.599g of Nickel chloride, and 136.315g of Zinc chloride. From these stock solutions, four distinct concentrations of each heavy metal salt solution were prepared, specifically at 25 mg/l, 50 mg/l, 75 mg/l, and 100 mg/l.

The water ferns Salvinia natans, S. molesta, and S. auriculata, all of similar size and quantity, were subjected to varying concentrations of different heavy metal salts in separate aquarium setups. These setups were left undisturbed in a shaded environment for a duration of 10 days. Following this treatment period, the plants were individually harvested from each container. They were subsequently rinsed with distilled water to eliminate any residual salts. Excess moisture was absorbed using tissue paper before the treated plants underwent analysis.

AS Analysis

Accurate weights of dried samples of Salvinia natans, Salvinia molesta, and Salvinia auriculata were obtained and subsequently dissolved in a mixture of HNO3 and HClO4 in a 3:1 ratio. The resulting solutions were evaporated to dryness and then extracted using distilled, deionized water. The solutions were brought to a boil and filtered. Each diluted sample was adjusted to a final volume of 100 mL. A 1.0 L water sample was heated to concentrate it, acidified, and then also adjusted to a total volume of 100 mL. The concentrations of metal ions in all samples were determined using an Atomic Absorption Spectrometer. Calibration charts for each metal were utilized to ascertain the unknown concentrations. All analyses were performed in triplicate, along with one blank sample.

The assessment of heavy metals was conducted utilizing the Atomic Absorption Spectrometer (AAS). The GBC 933A AAS model was employed for the quantification of heavy metals. Additionally, a Sargent-Welch digital pH meter, model Pax S-29998, featuring a glass electrode and capable of measuring pH with a precision of 0.01 within a range of 0 to 14, was utilized.

Conclusion

This process utilizes plants to extract heavy metals through a mechanism known as bioaccumulation. The current study has demonstrated that Salvinia species are highly invasive and can be effectively utilized for the phytoremediation of aquatic ecosystems contaminated with hazardous heavy metals, including Chromium, Copper, Zinc, Mercury, and Lead. The plants were subjected to various chemical compositions, and their phytoaccumulation capabilities were assessed. Characteristics such as high productivity, significant absorption capacity, and substantial metal removal potential position Salvinia species as macrophytes with considerable promise for application in phytoremediation technologies. The biomass that has been treated with metal can be safely smelted at a later stage. Additionally, it can be utilized to generate biofuels such as bioethanol and biomethanol through the activity of specific genetically modified microorganisms. This approach helps reduce pollution and safeguards aquatic ecosystems.

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