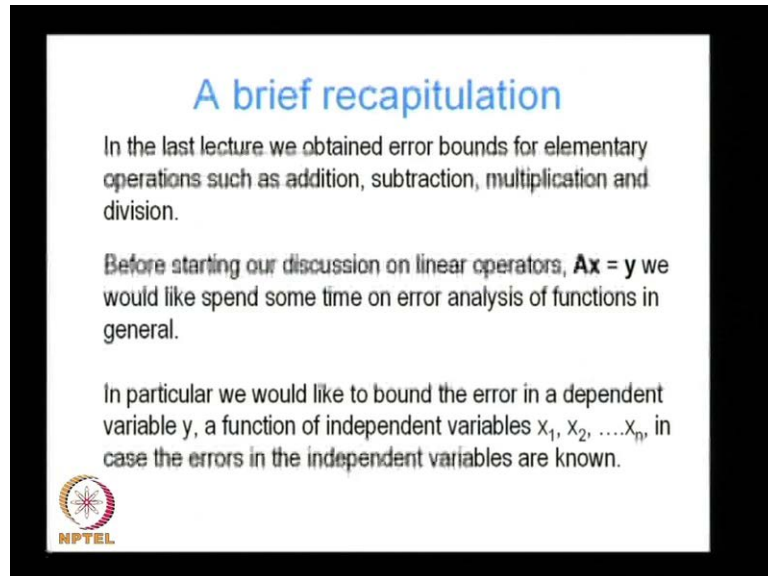


Numerical Methods in Civil Engineering
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Lecture - 3
Introduction to Linear Systems

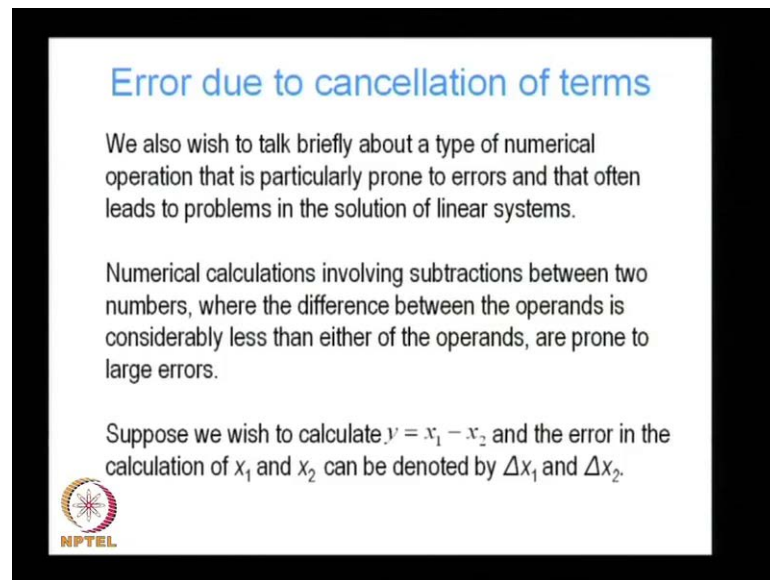
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In civil engineering, we are going to talk about Linear Systems. In the last lecture we obtained error bounds for elementary operations, such as addition subtraction, multiplication and division. Before starting our discussion on linear operators, which are typically written as A times x is equal to y , we would like to spend some time on error analysis of functions in general. In particular, we would like to bound the error in a dependent variable y , a function of in a dependent variables x_1, x_2 up to x_n .

In case the errors in the independent variables are known, for instance if our function is a linear function, if we know the errors in the depend in the independent variables $Ax = y$ then we want to find out what would be the error in y , after we operate on x operator with the linear operator A .

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Error due to cancellation of terms

We also wish to talk briefly about a type of numerical operation that is particularly prone to errors and that often leads to problems in the solution of linear systems.

Numerical calculations involving subtractions between two numbers, where the difference between the operands is considerably less than either of the operands, are prone to large errors.

Suppose we wish to calculate, $y = x_1 - x_2$ and the error in the calculation of x_1 and x_2 can be denoted by Δx_1 and Δx_2 .



We would also like to wish, we would also like to talk briefly about a type of numerical operation, that is particularly prone to errors, and that often leads to problems in the solution of linear systems. Numerical calculations involving subtractions between two numbers, where the difference between the operands is considerably less than either of the operands, are prone to large errors.

Suppose, we wish to calculate y is equal to x_1 minus x_2 and the error in the calculation of x_1 and x_2 can be denoted by Δx_1 and Δx_2 , that is x_1 and x_2 are my independent variables; and I know the error in x_1 and x_2 which is given by Δx_1 and Δx_2 .


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Error due to cancellation of terms

Then the error in y , Δy is given as:

$$|\Delta y| \leq |\Delta x_1| + |\Delta x_2|$$
$$\therefore \left| \frac{\Delta y}{y} \right| \leq \frac{|\Delta x_1| + |\Delta x_2|}{|x_1 - x_2|}$$

Then the relative error in y can be very large if $|x_1 - x_2|$ is small, i.e. the two numbers x_1 and x_2 are nearly equal



Then the error in y , Δy is given as bound on the error in Δy , mod of Δy is lesser than or equal to mod of Δx_1 plus mod of Δx_2 , we have seen this from our last lecture, where we found the bounds on simple operations like subtraction. So, from that equation dividing both sides by mod of y , we get Δy by y mod of that is lesser than or equal to mod of Δx_1 plus mod of Δx_2 divided by mod of x_1 minus x_2 . It is clear from the last expression, that the relative error in y will be very large, if mod of x_1 minus x_2 is small that is the two numbers x_1 and x_2 are nearly equal.

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Error due to cancellation of terms


For example, if

$x_1 = .5764$ and the absolute error in x_1 is $.5 \times 10^{-4}$
 $x_2 = .5763$ and the absolute error in x_2 is also $.5 \times 10^{-4}$

Then $x_1 - x_2 = .0001$ but the error bound on the numerical approximation of $x_1 - x_2$ is:

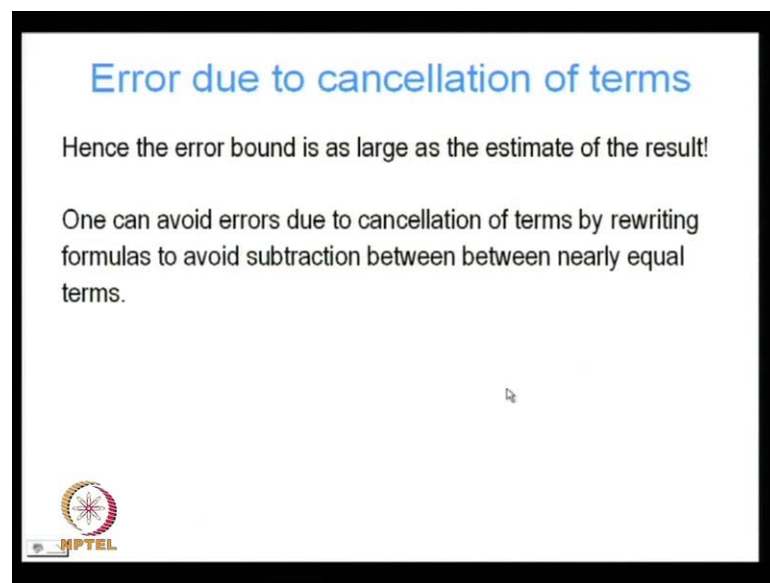
$$.5 \times 10^{-4} + .5 \times 10^{-4} = .001$$

Thus the error bound on $x_1 - x_2$ is as large as $x_1 - x_2$



For example, if x_1 is equal to 0.5764 and the absolute error in x_1 that is Δx_1 is 0.5×10^{-4} , and x_2 is equal to 0.5763, and the absolute error in x_2 is also 0.5×10^{-4} . Then $x_1 - x_2$ is equal to 0.0001, but the error bound on the numerical approximation of $x_1 - x_2$ is the sum of this absolute errors in x_1 and x_2 . So, that is equal to $0.5 \times 10^{-4} + 0.5 \times 10^{-4}$ equal to 0.001 , thus the error bound on $x_1 - x_2$ is as large as $x_1 - x_2$, recall $x_1 - x_2$ is 0.0001. But, the error bound is actually 0.001, which is actually in order of magnitude greater than $x_1 - x_2$.


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Error due to cancellation of terms

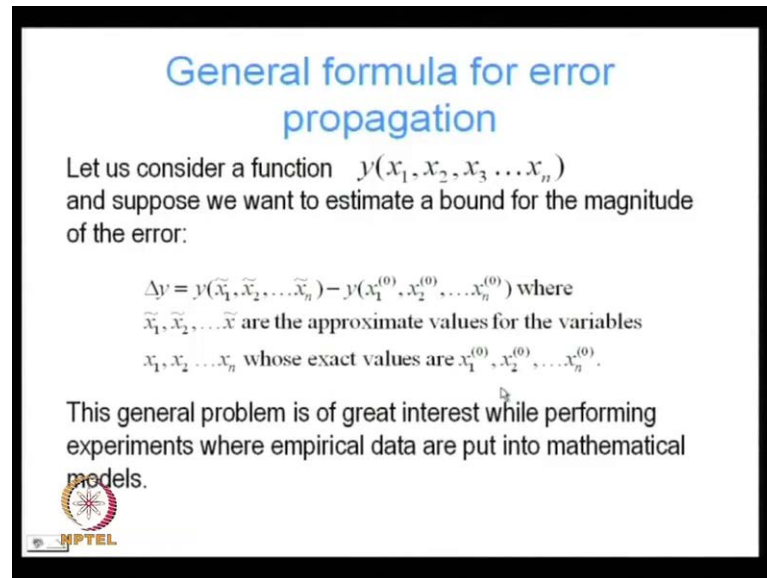
Hence the error bound is as large as the estimate of the result!

One can avoid errors due to cancellation of terms by rewriting formulas to avoid subtraction between nearly equal terms.



Hence, the error bound is as large as the estimate of the result, one can avoid errors due to cancellation of terms by rewriting formulas to avoid subtraction between nearly equal terms.

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General formula for error propagation

Let us consider a function $y(x_1, x_2, x_3, \dots, x_n)$ and suppose we want to estimate a bound for the magnitude of the error:

$$\Delta y = y(\tilde{x}_1, \tilde{x}_2, \dots, \tilde{x}_n) - y(x_1^{(0)}, x_2^{(0)}, \dots, x_n^{(0)})$$

where $\tilde{x}_1, \tilde{x}_2, \dots, \tilde{x}_n$ are the approximate values for the variables x_1, x_2, \dots, x_n whose exact values are $x_1^{(0)}, x_2^{(0)}, \dots, x_n^{(0)}$.

This general problem is of great interest while performing experiments where empirical data are put into mathematical models.



Let us consider in general a function y , which is in general function of x_1, x_2, x_3 up to x_n and let us suppose we want to estimate a bound for the magnitude of the error, that is we want to find out a bound on Δy . Where y is actually function of \tilde{x}_1, \tilde{x}_2 , up to \tilde{x}_n , minus y of $x_1^{(0)}, x_2^{(0)}$ up to $x_n^{(0)}$ where $\tilde{x}_1, \tilde{x}_2, \tilde{x}_n$ are the approximate values for the variables, while $x_1^{(0)}, x_2^{(0)}, x_n^{(0)}$ are the exact values of the variables.

So, the variables are x_1, x_2, x_n whose exact values are $x_1^{(0)}, x_2^{(0)}$ and $x_n^{(0)}$ and $\tilde{x}_1, \tilde{x}_2, \tilde{x}_n$ are the approximate values for those variables. We want to find given that we have the approximate values of x_1 through x_n , which is \tilde{x}_1 through \tilde{x}_n , we want to find what would be the error in y due to the error in x . This general problem is of great interest while performing experiments, where empirical data are put into mathematical models, for instance we do some experiments where we find the values of x_1 , where the independent variables in my experiments are x_1 through x_n .

I do that experiment and I find the values $\tilde{x}_1, \tilde{x}_2, \tilde{x}_n$, because of some experimental errors $\tilde{x}_1, \tilde{x}_2, \tilde{x}_n$ are not the exact values, which are actually $x_1^{(0)}, x_2^{(0)}$ and $x_n^{(0)}$. But, given \tilde{x}_1, \tilde{x}_2 and \tilde{x}_n up to \tilde{x}_n I want to find what is the error in y due to the error in my experiment, but due to my experimental error.


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General formula for error propagation

Denoting :

$$\tilde{x} = (\tilde{x}_1, \tilde{x}_2, \dots, \tilde{x}_n) \quad x_0 = (x_1^{(0)}, x_2^{(0)}, \dots, x_n^{(0)})$$
$$\Delta x_i = \tilde{x}_i - x_i^{(0)} \quad \Delta y = y(\tilde{x}) - y(x_0)$$

It can be shown that : $\Delta y \approx \sum_{i=1}^n \frac{\partial y}{\partial x_i}(\tilde{x}) \Delta x_i$ and hence

$$|\Delta y| \leq \sum_{i=1}^n \left| \frac{\partial y}{\partial x_i}(\tilde{x}) \right| |\Delta x_i|$$


So, we denote \tilde{x} is equal to x_1 tilde, x_2 tilde through x_n tilde and x_0 is equal to x_{10} , x_{20} through x_{n0} and we denote Δx_i is equal to x_1 tilde minus x_{i0} and Δy is equal to y of \tilde{x} minus y of x_0 . That is the numerical solution which is y of \tilde{x} minus the exact solution y of x_0 , which gives me my error in y , it can be shown the Δy is approximately equal to the sum of the partial derivatives of y , with respect to the independent variables x_i evaluated add the approximate solution \tilde{x} times Δx_i .

Since that is true the bound on y must be lesser than or equal to the sum of the individual partial derivatives, times the error in each of the individual variables, that is mod of Δy is lesser than or equal to $\sum_{i=1}^n \text{mod of } \frac{\partial y}{\partial x_i} \text{ evaluated at } \tilde{x} \text{ times mod of } \Delta x_i$. This result again we get from our previous lecture, where we talked about the error bounds on sums.

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General formula for error propagation

One can derive the above formula as follows:

Suppose we go from x_0 to \tilde{x} in n steps, changing one coordinate at a time. Thus for $i=1,2,\dots,n-1$, we set:

$$x_{i-1} = (\tilde{x}_1, \tilde{x}_2, \dots, x_i^{(0)}, x_{i+1}^{(0)}, x_{i+2}^{(0)}, \dots, x_n^{(0)})$$
$$\Downarrow$$
$$x_i = (\tilde{x}_1, \tilde{x}_2, \dots, \tilde{x}_i, x_{i+1}^{(0)}, x_{i+2}^{(0)}, \dots, x_n^{(0)})$$

Obviously,


$$x_n = \tilde{x}.$$

One can derive the previous formula, this formula as follows, suppose we go from x_0 to \tilde{x} , x_0 being the exact solution, \tilde{x} being the approximate solution in n steps. In each step we change we start with x_0, x_1, x_2, x_3 up to x_n at each step, we change one of the individual variables to its approximate value. For instance, when we go from step $i-1$ to step i , we change x_i to \tilde{x}_i , otherwise all the variables, all the independent variables have the same value between x_{i-1} and x_i .

It is only the independent variable x_i , we change its value from x_i to \tilde{x}_i obviously, if you perform this through n steps, at each step changing the value of one particular independent variable, at the end of the n steps x_n will be equal to our approximate solution \tilde{x} .

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General formula for error propagation


In going from the $i-1^{\text{th}}$ step to the i^{th} step, the i^{th} coordinate is changed from $x_i^{(0)}$ to \tilde{x}_i while the rest of the coordinates are left unchanged.

Then by the mean value theorem, there exists a $\xi_i \in [x_{i-1}, x_i]$ such that:

$$y(x_i) - y(x_{i-1}) = \frac{\partial y}{\partial x_i}(\xi_i) \cdot (\tilde{x}_i - x_i^{(0)})$$

$$\approx \frac{\partial y}{\partial x_i}(\tilde{x}) \cdot \Delta x_i \quad (*)$$

where the partial derivative of y w.r.t. x_i at ξ_i is approximated by the partial derivative at the end of n steps



In going from the $i-1^{\text{th}}$ step to i^{th} step, the i^{th} coordinate is changed from x_{i-1} to \tilde{x}_i , while the rest of the coordinates are left unchanged. Because of this, due to the mean value theorem, there exists a ξ_i which belongs to the interval $x_{i-1} \leq \xi_i \leq x_i$ such that, $y(x_i) - y(x_{i-1}) = \frac{\partial y}{\partial x_i}(\xi_i) \cdot (\tilde{x}_i - x_{i-1})$ is equal to $\frac{\partial y}{\partial x_i}$ evaluated at ξ_i times the interval which is equal to $\tilde{x}_i - x_{i-1}$, which is approximately equal to $\frac{\partial y}{\partial x_i}$ evaluated at \tilde{x} dotted with Δx_i . The approximation is because, we have evaluated the partial derivative of y with respect to x_i at ξ_i by the partial derivative at the end of my n steps, that is an evaluating $\frac{\partial y}{\partial x_i}$ at the final value which is \tilde{x} .

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General formula for error propagation

Since,

$$\Delta y = y(x_n) - y(x_0), \text{ we can write :}$$

$$\Delta y = \sum_{i=1}^n [y(x_i) - y(x_{i-1})]$$

Hence, (*) can be re-written as :

$$\Delta y \approx \sum_{i=1}^n \frac{\partial y}{\partial x_i}(\tilde{x}) \cdot \Delta x_i \leq \sum_{i=1}^n \left| \frac{\partial y}{\partial x_i}(\tilde{x}) \right| \cdot |\Delta x_i|$$


Since Δy is equal to y of x_n minus y of x_0 , recall y of x_0 is the solution where with the exact values for the independent variables. And y of x_n is the solution where we have replace the independent variables, the exact values of the independent variables with the approximate values. So, Δy is equal to y of x_n minus y of x_0 , we can actually write Δy as the sum over i to n y of x_i minus y of x_{i-1} , it turns out that if you do this summation all the intermediate quantities are going to cancel out. And we would be left with y of x_n with i equal to n minus y of x_0 were i is equal to 1.

Because of this, we can rewrite this expression as Δy is approximately equal to summation, basically we are taking the sum ((Refer Time: 12:44)), if we take the sum on both sides of this expression. Then we get Δy from this expression, Δy is approximately equal to $\sum_{i=1}^n \frac{\partial y}{\partial x_i} \tilde{x}_i \cdot \Delta x_i$, which again from our bounds an additions is lesser than or equal to $\sum_{i=1}^n \text{mod of } \frac{\partial y}{\partial x_i} \tilde{x}_i \cdot \text{mod of } \Delta x_i$.


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General formula for error propagation

In the above Δy was approximated with the total differential of y . This means that in a small nbd. of \tilde{x} , containing x_0 one can approximate y with a linear function.

Because of the approximation involved in evaluating the partial derivatives at \tilde{x} usually $\left| \frac{\partial y}{\partial x_i} \right|$ is evaluated at \tilde{x}

Then 5 to 10 percent of the value is added as a margin of safety.

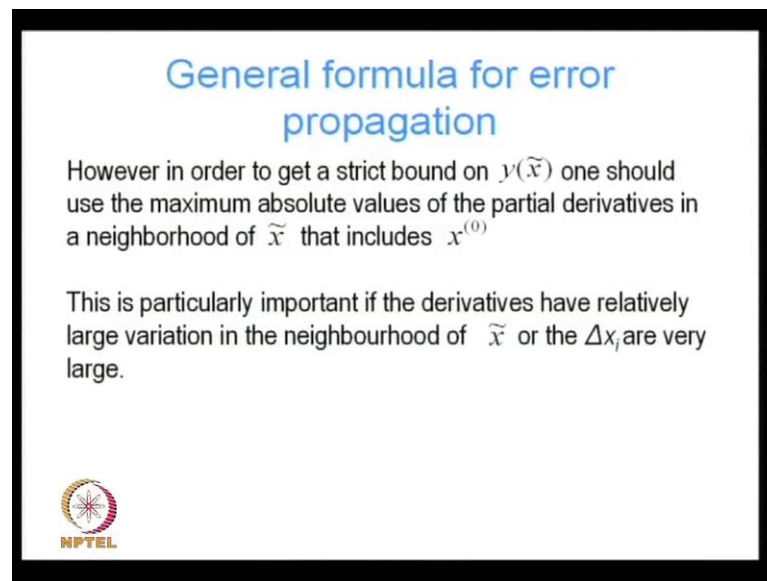


In the above if you go back to our previous equation and look, we are actually approximating Δy by the total differential of y , this means that in a small neighborhood of \tilde{x} containing x_0 , we are actually approximating y with a linear function. However, because of the approximation involved in evaluating the partial derivatives of at \tilde{x} , recall that instead of evaluating the partial derivatives at x_i , we

are evaluating the partial derivative at \tilde{x} , there is some approximation involved there.

So, what is actually done is that although we evaluate $\frac{\partial y}{\partial x_i}$ at \tilde{x} , 5 to 10 percent of the value is added as a margin of safety. Because, we are not evaluating $\frac{\partial y}{\partial x_i}$ at the exact point x_i , but at \tilde{x} , which is the approximate value.


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General formula for error propagation

However in order to get a strict bound on $y(\tilde{x})$ one should use the maximum absolute values of the partial derivatives in a neighborhood of \tilde{x} that includes $x^{(0)}$

This is particularly important if the derivatives have relatively large variation in the neighbourhood of \tilde{x} or the Δx_i are very large.



In order to get a strict bound of y of \tilde{x} , one should use the maximum absolute values of the partial derivatives in a neighborhood of \tilde{x} , that includes x_0 . This becomes particularly important, if the derivatives have relatively large variation in the neighborhood of \tilde{x} , or the Δx_i are very large. So, we are saying that in order to get a strict bound, we should not really evaluate the partial derivatives at \tilde{x} , we should actually look for the maximum value of the partial derivatives in a neighborhood of \tilde{x} , that includes the true solution x_0 .

This is important particularly when the derivatives are varying significantly, there is significant variation in the derivatives. So, it is important that way in order to obtain a proper upper bound, that we take the maximum value of the partial derivative in a neighborhood of \tilde{x} .


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General formula for error propagation

The bound for $|\Delta y|$ obtained above covers the worst possible cases where the sources of error Δx_i contribute with the same sign and cancellation of error contributions is not accounted for.

For this reason the bound for $|\Delta y|$ thus obtained is called the maximal error.

Often the maximal error bound, particularly for a large number of variables, is too coarse. It predicts unrealistically high error bounds.



The bound for mod of delta y obtained above, covers the worst possible cases where the sources of error delta x i contribute with the same sign, and cancellation of error contributions is not accounted for. For this reason the bound for mod of delta y thus obtained is called the maximal error, but the maximal error bound is often not very useful, particularly for a large number of variables, because it is true codes, it predicts an unrealistically high error estimate.

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A more realistic error bound

The theory of probability can be used to obtain a more realistic estimate of the error in terms of the standard deviations of the variables x and y

Assuming the x_i are independent random variables with mean zero and standard deviation ε_i , it can be shown that the standard deviation of y , ε is given by:

$$\varepsilon = \left[\sum_{i=1}^n \left(\frac{\partial y}{\partial x_i} \right)^2 \varepsilon_i^2 \right]^{\frac{1}{2}}$$

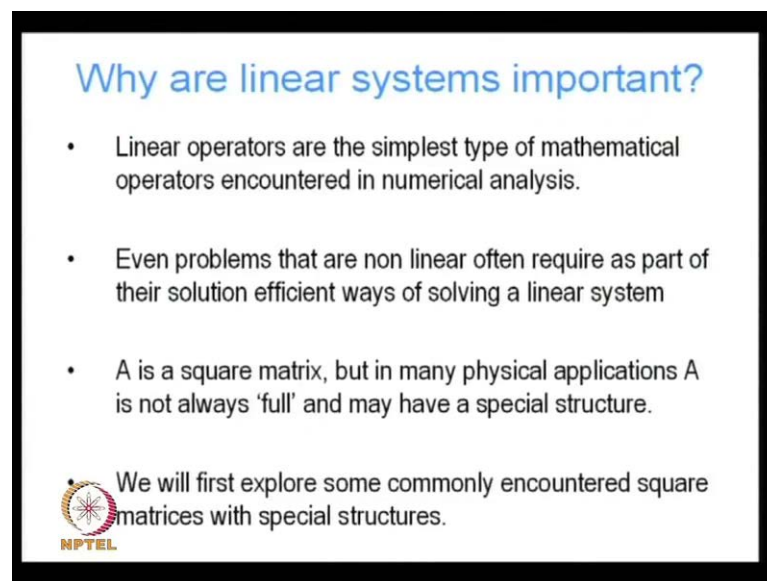
The standard deviation of y , ε is known as the standard error. It often gives a more realistic estimate of the true error than the maximal error.



So, instead of using the maximal error, sometimes we use something which is known as the standard error, how do we get the standard error, we use the theory of probability to obtain the standard error. What do we do, we assume that the x_i are independent random variables with mean 0 and standard deviation ϵ_i , if we assumed that it can be shown. That the standard deviation in the dependent variable y is given by a summation of the partial derivatives of y with respect to x_i square times the standard deviation square and the square root of that.

If that is ϵ , which is by standard deviation in y and this standard deviation is known as the standard error, it often gives a more realistic estimate of the true error than the maximal error. Because, it gives the narrower bound, if instead of the previous error bound, this error bound is much more realistic.


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Why are linear systems important?

- Linear operators are the simplest type of mathematical operators encountered in numerical analysis.
- Even problems that are non linear often require as part of their solution efficient ways of solving a linear system
- A is a square matrix, but in many physical applications A is not always 'full' and may have a special structure.

We will first explore some commonly encountered square matrices with special structures.

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So, with that discussion we are ready to talk about linear systems, why are linear systems important, because linear operators are basically the simplest type of mathematical operators encountered in numerical analysis. Even problems that are non-linear often require as to obtain solutions of linear systems, for instance if A is a square matrix, and many physical applications A is not always full, and may have a special structure. Recall our definition of linear systems which we obtained earlier, which says that we want to solve the system $Ax = y$. Typically in those linear systems A is a square matrix, but in many physical applications A is not always full, and may have a special structure.

We will first explore some commonly encountered square matrices with special structures.

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Triangular Matrices

Triangular Matrices

Triangular matrices have the following form:

$$L = \begin{bmatrix} l_{11} & 0 & \dots & 0 \\ l_{21} & l_{22} & 0 & \dots & 0 \\ \dots & \dots & \dots & \dots & \dots \\ l_{n1} & l_{n2} & l_{n3} & \dots & l_{nn} \end{bmatrix}$$


Lower triangular matrix

$$\det L = l_{11} \cdot l_{22} \cdot \dots \cdot l_{nn}$$

$$R = \begin{bmatrix} r_{11} & r_{12} & \dots & r_{1n} \\ 0 & r_{22} & r_{23} & \dots & r_{2n} \\ \dots & \dots & \dots & \dots & \dots \\ 0 & 0 & 0 & \dots & r_{nn} \end{bmatrix}$$

Upper triangular matrix

$$\det R = r_{11} \cdot r_{22} \cdot \dots \cdot r_{nn}$$



We will first talk about triangular matrices, triangular matrices have the following form, for instance we have some matrices known as lower triangular matrices, which contained only non-zero terms below or at the diagonal. Either the terms are non zero below the diagonal or along the diagonal, all terms above the diagonal are zero, that is a lower triangular matrix. And upper triangular matrix is a matrix which contains zeros along the diagonal, which contains zeros.

And all terms which are below the diagonal, which contain non zero terms along the diagonal and above the diagonal, that is an upper triangular matrix, but both upper and lower triangular matrix share a common property, which is that the determinant of such matrices are given by the products of the diagonal terms. So, for the lower triangular matrix, the determinant of L is equal to $l_{11} \cdot l_{22} \cdot \dots \cdot l_{nn}$, similarly for the upper triangular matrix the determinant of R is given by $r_{11} \cdot r_{22} \cdot \dots \cdot r_{nn}$.

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Partitioning Matrices

Partitioned Matrices

Often desirable to partition matrix into blocks of sub matrices

$$\begin{bmatrix} 7 & 3 & 9 & 2 \\ 4 & 5 & 6 & 11 \\ 13 & 2 & 14 & 19 \\ 2 & 5 & 19 & 13 \end{bmatrix} = \begin{bmatrix} 7 & | & 3 & 9 & 2 \\ 4 & | & 5 & 6 & 11 \\ 13 & | & 2 & 14 & 19 \\ 2 & | & 5 & 19 & 13 \end{bmatrix} = \begin{bmatrix} 7 & 3 & | & 9 & 2 \\ 4 & 5 & | & 6 & 11 \\ 13 & 2 & | & 14 & 19 \\ 2 & 5 & | & 19 & 13 \end{bmatrix}$$

In general, any matrix A can be thought of as being built of matrices of lower order:

$$A = \begin{bmatrix} A_{11} & A_{12} & \dots & A_{1n} \\ A_{21} & A_{22} & \dots & A_{2n} \\ \dots & \dots & \dots & \dots \\ A_{n1} & A_{n2} & \dots & A_{nn} \end{bmatrix}$$


Next the other type of typical matrix that we are going to encounter in the solution of linear systems, are partitioned matrices, it is often desirable to partition a matrix into blocks of sub matrices. In general any matrix can be thought of as being built of matrices of lower order, for instance if A is partition matrix comprising of A_{11} A_{12} through A_{1n} , A_{21} A_{22} through A_{2n} , A_{n1} A_{n2} through A_{nn} . So, these A_{11} A_{12} A_{1n} these are all actually matrices, but they are partitions of the bigger matrix A, so A_{11} A_{12} are all matrices of size smaller than the full matrix A.


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Operations on Partitions

- Addition and subtractions on partitions can be performed as if the blocks were scalars:

$$C_{ij} = A_{ij} + B_{ij} \quad \text{where } A_{ij} \text{ and } B_{ij} \text{ are blocks of the partitioned matrix } A \text{ and } B$$
- For the product $C = A.B$, C is a partitioned matrix with each of its several blocks obtained by summing the products of blocks of A and B with the same row index and same column index

$$C_{ij} = \sum_{k=1}^n A_{ik} B_{kj}$$



Operations on partitions addition and subtraction and partitions can be performed as if the blocks were scalars, that is if C_{ij} is the summation of two partition matrices A and B . Then C_{ij} is basically the sum of the partitions A_{ij} plus B_{ij} , A_{ij} and B_{ij} being blocks of the partition matrix A and B . So, basically let me go with that again if A and B are partition matrices with blocks A_{ij} and B_{ij} , then if we want to sum those matrices A and B to obtain a third matrix C . Then the blocks of C , C_{ij} are just the sum of the individual blocks A_{ij} and B_{ij} of A and B .

If we want to evaluate the product C of two matrices A and B , where A and B are partition matrices, then we obtain the product by summing the products of blocks of A and B with the same row index, and the same column index. That is if A_{ik} and B_{kj} are partitions of A and B , we obtain C_{ij} a partition of C by taking the product of $A_{ik} B_{kj}$ summing over k equal to 1 to n . So, by treating these individual blocks, as if there elements of a matrix we can obtain the partition C , corresponding to the product of the partition matrices A and B .

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Block Diagonal Matrices


A block diagonal matrix is a matrix that can be written in partitioned form as $A = \text{diag}(A_{11}, A_{22}, \dots, A_{nn})$ with the A_{ij} as square matrices:

$$A = \begin{bmatrix} A_{11} & & 0 \\ & A_{22} & \\ 0 & & \ddots \\ & & & A_{nn} \end{bmatrix}$$

Similarly it is possible to define block triangular matrices:

$$L = \begin{bmatrix} L_{11} & 0 & \dots & 0 \\ L_{21} & L_{22} & 0 & \dots & 0 \\ \dots & \dots & \dots & \dots & \dots \\ L_{n1} & L_{n2} & L_{n3} & \dots & L_{nn} \end{bmatrix} \quad R = \begin{bmatrix} R_{11} & R_{12} & \dots & R_{1n} \\ 0 & R_{22} & R_{23} & \dots & R_{2n} \\ \dots & \dots & \dots & \dots & \dots \\ 0 & 0 & 0 & \dots & R_{nn} \end{bmatrix}$$

$\det L = L_{11} \cdot L_{22} \cdot \dots \cdot L_{nn}$ $\det R = R_{11} \cdot R_{22} \cdot \dots \cdot R_{nn}$



Next we want to talk about block diagonal matrices, a block diagonal matrix is a matrix that can be written in partition form as A is equal to diagonal $A_{11} A_{22}$ through A_{nn} that is A is a partition matrix, but $A_{11} A_{22} A_{nn}$ are actually matrices, they are not individual terms in the matrix A , they may be full matrices. But, the occupy diagonals on my full matrix A , on my matrix A on the diagonals I instead of having individual terms, I

have actually matrices. So, this type of matrix is known as a block diagonal matrix, it is as if the diagonal elements of A are actually blocks of matrices, blocks partitions.

Similarly, it is possible to define block triangular matrices, which have the same structure is triangular matrices, but instead of individual entities on the diagonal or below the diagonal we actually have blocks of matrices, at these locations. For instance L_{11} L_{21} L_{22} are actually blocks of matrices, but they have with the same structure as a lower triangular matrix. Similarly, a block upper triangular matrix has the same structure, as an upper triangular matrix the only difference is this R_{11} R_{12} R_{1n} , instead of being individual scalars are actually individual matrices.

However, when we tried to the evaluate the determinant of this block triangular matrix L, we can basically do the same, we can do the same thing that we did earlier. So, determinant of L is equal to L_{11} times L_{22} times L_{nn} exactly like we did earlier, similarly determinant of R is equal to R_{11} times R_{22} through R_{nn} . So, the same structure follows, the same structure with individual upper triangular, lower triangular matrices had passes over continues into block triangular matrices.


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Vector spaces

- The set of all n dimensional vectors form a vector space R_n of dimension n
- The inner product of two vectors x, y is calculated by:

$$(x, y) = x_1y_1 + x_2y_2 + x_3y_3 + \dots + x_ny_n$$
- If x is considered to be a column vector, the inner product can be interpreted as a special case of matrix multiplication:

$$(x, y) = x^T y$$



Next we want to talk about vector spaces, the set of all n dimensional vectors form a vector space R_n of dimension n , for instance the set of all three dimensional vectors form a vector space R_3 of dimension three. Similarly, for two dimensions, the set of all two dimensional vectors form a vector space R_2 of dimension two. The inner product of


two vectors x and y is calculated as follows, how we are defining the inner product, we are saying that the inner product is defined by the product of the components summed together. So, the inner product of two vectors x and y is equal to $x_1 y_1$ plus $x_2 y_2$ plus $x_3 y_3$ through $x_n y_n$. If x is considered to be a column vector, the inner product can be interpreted in matrix notation, the inner product between x and y can be interpreted in matrix notation, as the transpose of x times y is the product of the transpose of x and y .

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Linear Independence

- The vector $y = c_1x_1 + c_2x_2 + \dots + c_nx_n$ is a linear combination of the vectors x_1, x_2, \dots, x_n
- The vectors x_1, x_2, \dots, x_n are said to be linearly dependent if there is some set of constants c_1, c_2, \dots, c_n not all equal to zero, such that:

$$y = c_1x_1 + c_2x_2 + \dots + c_nx_n = 0$$
- The vectors x_1, x_2, \dots, x_n are said to be linearly independent when $y=0$ if and only if c_1, c_2, \dots, c_n are identically equal to zero



Next we want to talk about the notion of linear independence, a vector y which is obtained by taking a linear combination of vectors x_1, x_2 through x_n by multiplying them with scalar C_1 through C_n , is said to be in linearly independent. If there is some set of constant C_1, C_2 through C_n not all equal to 0 such that, that sum $C_1 X_1$ plus $C_2 X_2$ through $C_n X_n$ is equal to 0. On the other hand, the vectors x_1, x_2 through x_n are said to be linearly independent, when this condition is only satisfied if and only if all the C_1 through C_n are 0.

So, x_1 through x_n are linearly dependent, if there is some set of constants C_1 through C_n not all equal to 0 such that, $C_1 X_1$ plus $C_2 X_2$ plus through $C_n X_n$ is equal to 0, X_1 through X_n is linearly independent if that condition is only satisfied if all the C_1 are 0. Basically we are saying there is no set of non zero constants C_1 through C_n such that, we can sum x_1, x_2 through x_n after multiplying them with the scalars C_1 through C_n such that, we get 0.


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Linear Independence

- The maximum number of linearly independent vectors in R_n is n
- Any set of linearly independent vectors x_1, x_2, \dots, x_n form a basis for the vector space R_n
- Any vector v belonging to R_n can be expressed as a linear combination of the basis vectors x_1, x_2, \dots, x_n i.e.

$$v = \alpha_1 x_1 + \alpha_2 x_2 + \dots + \alpha_n x_n$$

with not all $\alpha_i = 0$




The maximum number of linearly independent vectors in R_n is n , any set of linearly independent vectors x_1 through x_n form a basis for the vector space R_n . Any vector v belonging to R_n can be expressed as a linear combination of the basis vectors x_1 through x_n , that is v if x_1 through x_n comprise a basis for this space R_n . Any vector v can be represented as a linear combination of x_1 through x_n , what do you mean by a linear combination, we can write it as a scalar multiple α_1 times x_1 plus α_2 times x_2 plus α_3 times x_3 through α_n times x_n , where not all the α_i are equal to 0.

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Linear Subspaces

- The set of vectors given by all possible linear combinations of vectors x_1, x_2, \dots, x_k with $k < n$, form a linear subspace R .
- The vectors x_1, x_2, \dots, x_k span the subspace R .
- The subspace R has dimension k if x_1, x_2, \dots, x_k are linearly independent.




Next we want to talk about linear subspaces, the set of vectors given by all possible linear combinations of the vectors x_1, x_2 through x_k with k less than n form a linear subspace R . And the vectors x_1, x_2 through x_k are set to span the subspace R , the subspace R has dimension k , if x_1, x_2 through x_k are linearly independent, that is the subspace R has got basis vectors x_1, x_2 through x_k .

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Rank of a matrix

- A matrix $A (m, n)$ can be thought of as being built up of column vectors and row vectors.
- The maximum number of linearly independent column vectors of A is equal to the maximum number of independent row vectors of A and called the rank of A
- Hence $\text{rank}(A) = r \leq \min(m, n)$
- If $r = m = n$, A is non singular i.e. it has a non-zero determinant and can be inverted.



Next we want to talk about the rank of A matrix, a rank of a matrix A is set to be r , if r is the maximum number of linearly independent column vectors or row vectors of A . So, basically the rank is the maximum number of linearly independent column vectors or row vectors of A ; and those the maximum number of linearly independent column vectors of A is always equal to the maximum number of independent row vectors of A . Rank of A therefore, must be lesser than or equal to the smaller of m and n , where m is the number of rows in A , and n is the number of columns in A .


Since, the number of linearly independent columns must be equal to the number of linearly independent rows, it is obvious that the rank which is equal to the number of linearly independent rows, and number of linearly independent columns must be lesser than or equal to the minimum of m and n . If the rank is equal to m is equal to n that is if m is equal to n means A is a square matrix, and if the rank is equal to m and equal to n , then the matrix A is said to be nonsingular that is it has got a non zero determinant, that is it can be invertible.

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Linear system of equations

- A linear system of equations with m equations and n unknowns:
$$a_{11}x_1 + a_{12}x_2 + \dots + a_{1n}x_n = b_1$$
$$a_{21}x_1 + a_{22}x_2 + \dots + a_{2n}x_n = b_2$$
$$\dots$$
$$a_{m1}x_1 + a_{m2}x_2 + \dots + a_{mn}x_n = b_m$$

can be written in matrix notation as $Ax = b$




A linear system of equations with m equations and n unknowns, we will be actually write it out, which we have in done up till now can be written like this $a_{11} a_{12} \dots a_{1n}$ are the components of the co-efficient matrix A . x_1, x_2 through x_n are my independent variables x , I can think of x as a vector with of dimension n and b is my right hand side, and we want to solve for x_1, x_2 through x_n with satisfies this condition Ax equal to b .

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Homogenous system

- If $b = 0$, the system is said to be homogeneous
- A homogenous system always has a trivial solution, $x = 0$
- If $\text{rank}(A) = r < n$, then $Ax = 0$, has $(n-r)$ independent solutions
- Any vector x which satisfied $Ax = 0$ is a "null" vector of A and the set of all such vectors comprise the null space of A .
- Since $Ax = 0$, has $(n-r)$ independent solutions, the null space of A has dimension $(n-r)$



If b is equal to 0 , the system is said to be homogeneous, so Ax is equal to 0 is said to be a homogeneous system, a homogeneous system always has a trivial solution, what do we

mean by trivial solution, that is a homogeneous system will always be satisfied by x_1, x_2, x_3 to through x_n equal to 0. If rank of A is equal to r is less than n , then Ax equal to 0 has n minus r independent solutions, any vector x which satisfies Ax is equal to 0 is a null vector of A .


And the set of all such vectors comprise the null space of A , the null space of A has dimension n minus r , because as we recall the set that if rank of A is equal to r less than n , then Ax is equal to 0 has n minus r independent solutions, so the null space of A has dimension n minus r .

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Existence of solutions

- The system of equations $Ax = b$ can be written as

$$x_1a_1 + x_2a_2 + \dots + x_na_n = b$$
 where $a_1, a_2 \dots$ are the columns of A
- Thus b can be expressed as a linear combination of the columns of A
- Hence for the system $Ax = b$ to have a solution, b must belong to the subspace R spanned by the columns of A



The system of equations Ax is equal to b can be written as $x_1 a_1$ plus $x_2 a_2$ through $x_n a_n$ equal to b , basically we are taking the rows of A . And multiplying scaling each row by x_1 through x_n scaling the first row with the x_1 , the second row with the x_2 , the third row with the x_3 and so on, and so forth. And the last row a_n by x_n and setting that equal to b , this is just another way of writing the system Ax equal to b , but this way of writing makes clear that b is actually a linear combination of the columns of A , which again makes it obvious that if the system Ax is equal to b is to have a solution.


b must belong to the subspace r spanned by the columns of A , because we call that $x_1 a_1$ plus $x_2 a_2$ plus $x_n a_n$ is equal to b , which basically means that b is a linear combination of $a_1 a_2$ through a_n , x_1 and x_2 through x_n are just the scalar multiples. So, b is a linear combination of the columns of A , $a_1 a_2$ through a_n thus b must belong

to the subspace R which is spanned by the columns of A , so in this case columns of A are the basis of that subspace. And b is obtained by taking a linear combination of those basis vectors a_1, a_2 through a_n , after scaling them with the scalars x_1, x_2 through x_n .

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Existence of solutions

- This condition can also be expressed as $\text{rank}(A, b) = \text{rank}(A)$
- Since b is a linear combination of the columns of A (for it to be a soln. of $Ax = b$) the augmented system has the same rank as A
- If $m = n = r$ then system $Ax = b$ has a soln. for any vector b since now that A has full rank, the cols. of A form a basis for the space R_n .
- Thus any vector b can be expressed as a linear combination of the column vectors of A

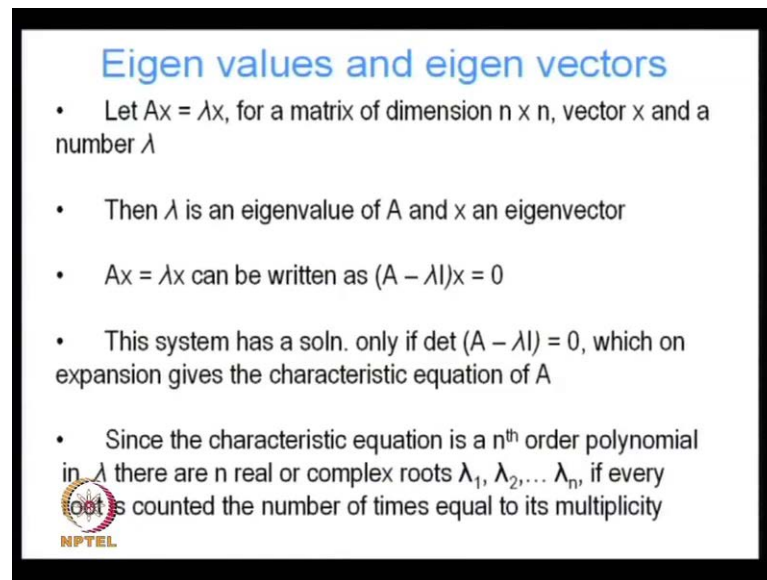


This condition can also be expressed by the requirement that rank of A, b is equal to rank of A , for this means since b is a linear combination of the columns of A that means, if we add we if we augment A by adding an additional column b we do not change the rank of A . Because, b is a linear combination of the columns of A , so we add an additional column to A which is b , it does not change the number of linearly independent columns of A . So, since b is a linear combination of the columns of A , the augmented system has the same rank as A .

If m is equal to n is equal to r that is if A is square matrix, and as got full rank, then $Ax = b$ has a solution for any vector b , since now that A has full rank, the columns of A form a basis for the space R_n . A has full rank means, A has rank n that means, each of the columns of A are linearly independent vectors and since, the space has dimension n and since I have n linearly independent vectors, which are the columns of A , which means that this implies that the columns of A form a basis for the space R_n .

And since the form a basis for the space of R_n , any vector b can be represented by a linear combination of the column vectors of A , thus any vector b can be solved for, even any vector b we can find x such that, $Ax = b$.

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Eigen values and eigen vectors

- Let $Ax = \lambda x$, for a matrix of dimension $n \times n$, vector x and a number λ
- Then λ is an eigenvalue of A and x an eigenvector
- $Ax = \lambda x$ can be written as $(A - \lambda I)x = 0$
- This system has a soln. only if $\det(A - \lambda I) = 0$, which on expansion gives the characteristic equation of A
- Since the characteristic equation is a n^{th} order polynomial in λ there are n real or complex roots $\lambda_1, \lambda_2, \dots, \lambda_n$, if every root is counted the number of times equal to its multiplicity

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Next we wish to talk about Eigen values and eigen vectors, let $Ax = \lambda x$, for a matrix of dimension n by n , a vector x and a scalar λ . Then λ is an Eigen value of A and x is an eigen vector of A , we can rewrite $Ax = \lambda x$ as $(A - \lambda I)x = 0$, where I is the identity matrix. $(A - \lambda I)x = 0$ is a homogeneous system, from what we have looked at previously, this homogeneous system is only going to have a solution, if determinant of $(A - \lambda I)$ is equal to 0.

If we expand this determinant out, if A write out the determinant, determinant $(A - \lambda I)$ we get a polynomial in λ , and that polynomial in λ is called the characteristic equation of A . And depending on the dimension of A , the dimension of the polynomial also varies for instance, if A is a 3 by 3 matrix the characteristic equation is a polynomial of dimension three. If A is the 4 by 4 matrix the characteristic equation is a polynomial of dimension four, and depending on the roots of the polynomial we get the Eigen values of A , $\lambda_1, \lambda_2, \dots, \lambda_n$ which may be real or complex numbers.

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Similarity transformations

- Let C be a non-singular matrix. Then the transformation $C^{-1}AC$ is said to be similar to A , and the transformation is called a similarity transformation.
- Then if $Ax = \lambda x$, $(C^{-1}AC) C^{-1}x = \lambda C^{-1}x$
- Denoting $B = C^{-1}AC$, we have $B(C^{-1}x) = \lambda(C^{-1}x)$
- Thus after similarity transformation the eigen values are preserved while the eigen vectors are transformed as $C^{-1}x$



But, if we count each root equal to the number of times equal to it is multiplicity, we get n real or complex roots of that polynomial which are the Eigen values of A . Next, we want to talk about similarity transformations, let C be a non singular matrix that is C has determinant which is non zero, that is determinant of C is greater than 0. Then the transformation $C^{-1}AC$ is said to be similar to A , that is if we start with the original matrix A .

And when we transform it according to the rules $C^{-1}AC$, then the matrix we get $C^{-1}AC$ is said to be similar to A , and the transformation is called a similarity transformation. Let us go back to the Eigen value problem $Ax = \lambda x$ and suppose, we perform a similarity transformation on A to get $C^{-1}AC$, we can see that if we pre multiply both the left hand and right hand side with the matrix C^{-1} , we get $C^{-1}Ax = \lambda C^{-1}x$ which is equal to $\lambda C^{-1}x$.

I can rewrite as $C^{-1}AC$ operating on $C^{-1}x$, because C and C^{-1} on taking the product they give me the identity matrix. So, I can rewrite $Ax = \lambda x$ as $C^{-1}AC$ operating on $C^{-1}x$, which is equal to $\lambda C^{-1}x$. But, in this last equation, this has the exact structure of an Eigen value problem which becomes evident, if we replace $C^{-1}AC$ with B , now so that we can write $B(C^{-1}x) = \lambda(C^{-1}x)$.

Thus we see that after similarity transformation of A, the similar matrix C inverse A C has the same eigen values of as A, it has a same Eigen values lambda, but it is eigenvectors are no longer x. They are now C inverse x, thus after similarity transformation the eigen values are preserved, where the eigenvectors are transformed as C inverse of x.


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Similarity transformations

- An arbitrary square matrix of dimension $n \times n$ has n eigen values and n eigen vectors, $Ax_i = \lambda x_i, i = 1, \dots, n$
- These n equations can be written as:

$$AX = \Lambda X \quad \text{where } X = (x_1, x_2, \dots, x_n)$$

$$\Lambda = \text{diag}(\lambda_1, \lambda_2, \dots, \lambda_n)$$
- If the eigen vectors are linearly independent, then X has full rank and is non singular.
- Hence we can write: $X^{-1}AX = \Lambda$



An arbitrary square matrix of dimension n by n has n eigen values and n eigen vectors, each of which satisfied the relation $A x_i$ is equal to λx_i , these n equations. If we can combined all these equations together, we can write it in matrix form as $A x$ is equal to λx . Where now $A x$ is no longer of vector x is a combination of all this eigen vectors each eigen vector occupies a column of the capital matrix x . So, x_1, x_2 through x_n are individual eigen vectors and have combined them together, to form the matrix big x and λ is another matrix, whose which is basically a diagonal matrix.


And the diagonal elements of λ are the individual eigen values λ_1, λ_2 through λ_n corresponding to the eigen vectors x_1, x_2 through x_n . If the eigen vectors are linearly independent X is going to have full rank, and is going to be non-singular why, because if the eigenvectors are linearly independent each of the columns of A are linearly independent. And each of the columns of x are linearly independent, x has full rank and that is x can be invertible and x is non-singular. Since, x is invertible we can write $x^{-1} A x$ is equal to λ , basically we scale, we pre multiply $A x$ is

equal to λx with x inverse both on the left hand and the right hand side, when we get x inverse $A x$ is equal to λx . And recall λ is a diagonal matrix that is it has got 0 is everywhere, and $\lambda_1 \lambda_2$ through λ_n on the diagonals.

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Similarity transformations

- Thus a similarity transformation by X , the matrix of eigen vectors, transforms A into a diagonal matrix which has the eigen values of A as its diagonal terms.
- If the eigen values of A are distinct, i.e. λ_i is not equal to λ_j for all i not equal to j , then the eigen vectors are always linearly independent.
- For a real symmetric matrix, the eigen vectors corresponding to distinct eigen values are in addition orthogonal i.e. $(x_i)^T x_j = 0$




Thus a similarity transformation by X , the matrix of eigenvectors recall that earlier the transformation that we look that, this relationship this is basically a similarity transformation on A . So, what we are saying is that, if we do a similarity transformation on A using the matrix X , we get a diagonal matrix we transform A into a diagonal matrix which has the Eigen values of A as it is diagonal terms. If the Eigen values of A are distinct that is λ_i is not equal to λ_j for any i or j , then the eigen vectors are always going to be linearly independent.

So, if a matrix has got distinct eigen values that each of it is Eigen values are different from each other, then the eigenvectors are also going to be linearly independent. And we know that for a real symmetric matrix, the eigen vectors corresponding to distinct Eigen values, in addition to being linearly independent they are also orthogonal. So, that is not really true for always, but it is only true for real symmetric matrices, when the eigenvectors corresponding to distinct eigen values are mutually orthogonal. What it means, what does orthogonal mean it means that, if we take the dot product of two of those eigen vectors x_i dotted with x_j or been matrix notation $x_i^T x_j$, we are going to get 0.

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Similarity transformations

- Even if all the eigen values are not distinct, the eigen vectors of a symmetric matrix, corresponding to the same eigen values, can be chosen to be orthogonal.
- Thus for every symmetric matrix there is an orthogonal matrix X such that $X^T X = I$, such that $\Lambda = X^T A X$
- $(A + cI)x = (\lambda + c)x$ i.e. adding constant diagonal terms to a matrix, shifts its eigenvalues by a constant
- $A^2 x = A \cdot Ax = A \lambda x = \lambda^2 x$. Hence, $A^n x = \lambda^n x$

 Thus if $P(z) = a_0 z^n + a_1 z^{n-1} + \dots + a_n$ be an arbitrary polynomial then $P(A)$ has eigen values $P(\lambda)$

Even if all the eigen values are not distinct, the eigenvectors of a symmetric matrix corresponding to the same eigen values can be chosen to be orthogonal, there are procedures for orthogonalization, which we are going to talk about later in this course. For basically for a symmetric matrix corresponding to distinct eigen values, corresponding to an eigen value, if there is more than one eigen vector we can make those eigenvectors orthogonal to each other.

We can define the eigen vectors, so that dot product to those eigenvectors is equal to 0, thus for every symmetric matrix, there is an orthogonal matrix x such that, $X^T x$ is equal to i that is such that, λ is equal to $X^T A x$. Next, we want to see how the eigen values change if we perform certain simple operations on a matrix, for instance we have an initial matrix A . And suppose we add the same constant scalar C to each of the diagonal terms of A , that is we form the matrix A plus $C i$, i being the identity matrix.

In that case what happens is that the eigen values of A just get shifted, that is instead of the eigen values remaining as λ for the eigen values become $\lambda + C$. So, adding a constant to the diagonal term just shifts the eigen value by a constant C , so that is one property of eigen values the second property, we want to talk about is. If we take powers of a matrix for instance, if we take a product of a matrix with itself, for instance A^2 I can write it as A product with $A x$, but $A x$ I know I can write as λx .


Thus A operating on $A x$ can write as A operating on λx , but we can since λ is a scalar I can move it out and so we have λ times $A \lambda x$ is equal to λ times $A x$, but again $A x$ is again equal to λx . So, we are going to get $\lambda^2 x$, thus if we take a power of matrix, if we take n th power of A , then the eigen value of the n th power of A , is basically the eigen value of A raised to the n th power.

So, the eigen value of A to the power n is equal to λ to the power n , because of this property if P is a polynomial of the form $A^0 z$ to the power n plus $A^1 z$ to the power n minus 1 through A^n , an instead of z if we replace that by the matrix A , then we can see that $P A$ is going to have eigen values P of λ .

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Singular Value Decomposition

- Let A be a $m \times n$ matrix of rank r . Then there exist:
- A $m \times m$ orthogonal matrix U (i.e. $U^T U = I$)
- A $n \times n$ orthogonal matrix V (i.e. $V^T V = I$)
- A $r \times r$ diagonal matrix D with strictly positive elements, called the singular values of A such that:

$$A = U \Sigma V^T, \quad \Sigma = \begin{bmatrix} D & 0 \\ 0 & D \end{bmatrix}$$


Finally, for the today's lecture I want to briefly introduce something, which is known as singular value decomposition. In our next lecture we are going to talk in much greater detail about this, but I just want to introduce this, so suppose A is an m by n matrix of rank R , then there will always be m by m orthogonal matrix recall what is an orthogonal matrix. An orthogonal matrix is a matrix all of whose columns are orthogonal to each other, so if I take the dot product of a column i , with a dot product of another column j I am going to get 0.

So, what I am saying is that, if we can always do a singular value decomposition of a matrix A into orthogonal matrices U and V and an r by r diagonal matrix D such that, I

can represent A as a product of U , Σ , and V^T . Σ is such that D occupies the diagonal elements of Σ .

Thank you, we will continue with this discussion in our next lecture.