# Mathematics for Machine Learning 

Adapted from the slides of<br>Dr. Alex Vakanski at U. of Idaho

## Notation

- $a, b, c$
- $\mathbf{x}, \mathbf{y}, \mathbf{z}$
- A, B, C
- A, B, C
- $X, Y, Z$
- $a \in \mathcal{A}$
- $|\mathcal{A}|$
- $\|\mathbf{v}\|$
- u $\cdot \mathbf{v}$ or $\langle\mathbf{u}, \mathbf{v}\rangle$
- $\mathbb{R}$
- $\mathbb{R}^{n}$
- $y=f(x)$ or $x \mapsto f(x)$
- $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$

Scalar (integer or real)
Vector (bold-font, lower case)
Matrix (bold-font, upper-case)
Tensor ((bold-font, upper-case)
Random variable (normal font, upper-case)
Set membership: $a$ is member of set $\mathcal{A}$
Cardinality: number of items in set $\mathcal{A}$
Norm of vector $\mathbf{v}$
Dot product of vectors $\mathbf{u}$ and $\mathbf{v}$
Set of real numbers
Real numbers space of dimension $n$
Function (map): assign a unique value $f(x)$ to each input value $x$
Function (map): map an $n$-dimensional vector into a scalar

## Notation

- $\mathbf{A} \odot \mathbf{B}$
- $\mathbf{A}^{\dagger}$
- $\frac{d^{n} f}{d x^{n}}$
- $\nabla_{\mathbf{x}} f(\mathbf{x})$
- $\mathbf{H}_{f}$
- $X \sim P$
- $P(X \mid Y)$
- $\mathcal{N}\left(\mu, \sigma^{2}\right)$
- $\mathbb{E}_{X \sim P}[f(X)]$
- $\operatorname{Var}(f(X))$
- $\operatorname{Cov}(f(X), g(Y))$
- $\operatorname{corr}(X, Y)$
- $D_{K L}(P \| Q)$
- $C E(P, Q)$

Element-wise product of matrices $\mathbf{A}$ and $\mathbf{B}$
Pseudo-inverse of matrix A
$n$-th derivative of function $f$ with respect to $x$
Gradient of function $f$ with respect to $\mathbf{x}$
Hessian matrix of function $f$
Random variable $X$ has distribution $P$
Probability of $X$ given $Y$
Gaussian distribution with mean $\mu$ and variance $\sigma^{2}$
Expectation of $f(X)$ with respect to $P(X)$
Variance of $f(X)$
Covariance of $f(X)$ and $g(Y)$
Correlation coefficient for $X$ and $Y$
Kullback-Leibler divergence for distributions $P$ and $Q$
Cross-entropy for distributions $P$ and $Q$

## Vectors

## Vectors

- Vector definition
- Computer science: vector is a one-dimensional array of ordered real-valued scalars
- Mathematics: vector is a quantity possessing both magnitude and direction, represented by an arrow indicating the direction, and the length of which is proportional to the magnitude
- Vectors are written in column form or in row form
- Denoted by bold-font lower-case letters

$$
\mathbf{x}=\left[\begin{array}{l}
1 \\
7 \\
0 \\
1
\end{array}\right] \quad \mathbf{x}=\left[\begin{array}{llll}
1 & 7 & 0 & 1
\end{array}\right]^{T}
$$

- For a general form vector with $n$ elements, the vector lies in the $n$-dimensional space $\mathbf{x} \in \mathbb{R}^{n}$

$$
\mathbf{x}=\left[\begin{array}{c}
x_{1} \\
x_{2} \\
\vdots \\
x_{n}
\end{array}\right]
$$

## Geometry of Vectors

## Vectors

- First interpretation of a vector: point in space
- E.g., in 2D we can visualize the data points with respect to a coordinate origin

- Second interpretation of a vector: direction in space
- E.g., the vector $\overrightarrow{\mathbf{v}}=[3,2]^{T}$ has a direction of 3 steps to the right and 2 steps up
- The notation $\overrightarrow{\mathbf{v}}$ is sometimes used to indicate that the vectors have a direction

- All vectors in the figure have the same direction
- Vector addition
- We add the coordinates, and follow the directions given by the two vectors that are added



## Geometry of Vectors

## Vectors

- The geometric interpretation of vectors as points in space allow us to consider a training set of input examples in ML as a collection of points in space
- Hence, classification can be viewed as discovering how to separate two clusters of points belonging to different classes (left picture)
- Rather than distinguishing images containing cars, planes, buildings, for example
- Or, it can help to visualize zero-centering and normalization of training data (right picture)



## Dot Product and Angles

## Vectors

- Dot product of vectors, $\mathbf{u} \cdot \mathbf{v}=\mathbf{u}^{T} \mathbf{v}=\sum_{i} u_{i} \cdot v_{i}$
- It is also referred to as inner product, or scalar product of vectors
- The dot product $\mathbf{u} \cdot \mathbf{v}$ is also often denoted by $\langle\mathbf{u}, \mathbf{v}\rangle$
- The dot product is a symmetric operation, $\mathbf{u} \cdot \mathbf{v}=\mathbf{u}^{T} \mathbf{v}=\mathbf{v}^{T} \mathbf{u}=\mathbf{v} \cdot \mathbf{u}$
- Geometric interpretation of a dot product: angle between two vectors
- I.e., dot product $\mathbf{v} \cdot \mathbf{w}$ over the norms of the vectors is $\cos (\theta)$

$$
\mathbf{u} \cdot \mathbf{v}=\|\mathbf{u}\|\|\mathbf{v}\| \cos (\theta) \quad \cos \theta=\frac{\mathbf{u} \cdot \mathbf{v}}{\|\mathbf{u}\|\|\mathbf{v}\|}
$$



- If two vectors are orthogonal: $\theta=90^{\circ}$, i.e., $\cos (\theta)=0$, then $\mathbf{u} \cdot \mathbf{v}=0$
- Also, in ML the term $\cos \theta=\frac{\mathbf{u} \cdot \mathbf{v}}{\|\mathbf{u}\|\|\mathbf{v}\|}$ is sometimes employed as a measure of closeness of two vectors/data instances, and it is referred to as cosine similarity


## Norm of a Vector

## Vectors

- A vector norm is a function that maps a vector to a scalar value
- The norm is a measure of the size of the vector
- The norm $f$ should satisfy the following properties:
- Scaling: $f(\alpha \mathbf{x})=|\alpha| f(\mathbf{x})$
- Triangle inequality: $f(\mathbf{x}+\mathbf{y}) \leq f(\mathbf{x})+f(\mathbf{y})$
- Must be non-negative: $f(\mathbf{x}) \geq 0$
- The general $\ell_{p}$ norm of a vector $\mathbf{x}$ is obtained as: $\|\mathbf{x}\|_{p}=\left(\sum_{i=1}^{n}\left|x_{i}\right|^{p}\right)^{\frac{1}{p}}$
- On next page we will review the most common norms, obtained for $p=1,2$, and $\infty$


## Norm of a Vector

## Vectors

- For $p=2$, we have $\ell_{2}$ norm
- Also called Euclidean norm
- It is the most often used norm

$$
\|\mathbf{x}\|_{2}=\sqrt{\sum_{i=1}^{n} x_{i}^{2}}=\sqrt{\mathbf{x}^{T} \mathbf{x}}
$$

- $\ell_{2}$ norm is often denoted just as $\|\mathbf{x}\|$ with the subscript 2 omitted
- For $p=1$, we have $\ell_{1}$ norm
- Uses the absolute values of the elements

$$
\|\mathbf{x}\|_{1}=\sum_{i=1}^{n}\left|x_{i}\right|
$$

- Discriminate between zero and non-zero elements
- For $p=\infty$, we have $\ell_{\infty}$ norm
- Known as infinity norm, or max norm
- Outputs the absolute value of the largest element
- $\ell_{0}$ norm outputs the number of non-zero elements
- It is not an $\ell_{p}$ norm, and it is not really a norm function either (it is incorrectly called a norm)


## Vector Projection

## Vectors

- Orthogonal projection of a vector $\mathbf{y}$ onto vector $\mathbf{x}$
- The projection can take place in any space of dimensionality $\geq 2$
- The unit vector in the direction of $\mathbf{x}$ is $\frac{\mathbf{x}}{\|\mathbf{x}\|}$
- A unit vector has norm equal to 1
- The length of the projection of $\mathbf{y}$ onto $\mathbf{x}$ is $\|\mathbf{y}\| \cdot \cos (\theta)$

- The orthogonal project is the vector $\operatorname{proj}_{\mathbf{x}}(\mathbf{y})$

$$
\operatorname{proj}_{\mathbf{x}}(\mathbf{y})=\frac{\mathbf{x} \cdot\|\mathbf{y}\| \cdot \cos (\theta)}{\|\mathbf{x}\|}
$$

## HyYDerolanes

## Hyperplanes

- Hyperplane is a subspace whose dimension is one less than that of its ambient space
- In a 2D space, a hyperplane is a straight line (i.e., 1D)
- In a 3D, a hyperplane is a plane (i.e., 2D)
- In a $d$-dimensional vector space, a hyperplane has $d-1$ dimensions, and divides the space into two half-spaces
- Hyperplane is a generalization of a concept of plane in high-dimensional space
- In ML, hyperplanes are decision boundaries used for linear classification
- Data points falling on either sides of the hyperplane are attributed to different classes



## Hyperplanes

## Hyperplanes

- For example, for a given data point $\mathbf{w}=[2,1]^{T}$, we can use dot-product to find the hyperplane for which $\mathbf{w} \cdot \mathbf{v}=1$
- I.e., all vectors with $\mathbf{w} \cdot \mathbf{v}>1$ can be classified as one class, and all vectors with $\mathbf{w} \cdot \mathbf{v}<1$ can be classified as another class

- Solving $\mathbf{w} \cdot \mathbf{v}=1$, we obtain

$$
\|\mathbf{v}\|\|\mathbf{w}\| \cos (\theta)=1 \Longleftrightarrow\|\mathbf{v}\| \cos (\theta)=\frac{1}{\|\mathbf{w}\|}=\frac{1}{\sqrt{5}}
$$

- I.e., the solution is the set of points for which $\mathbf{w} \cdot \mathbf{v}=1$ meaning the points lay on the line that is orthogonal to the vector $\mathbf{w}$
- That is the line $2 x+y=1$
- The orthogonal projection of $\mathbf{v}$ onto $\mathbf{w}$ is $\|\mathbf{v}\| \cos (\theta)=\frac{1}{\sqrt{5}}$



## HyYerolanes

Hyperplanes

- In a 3D space, if we have a vector $\mathbf{w}=[1,2,3]^{T}$ and try to find all points that satisfy $\mathbf{w} \cdot \mathbf{v}=1$, we can obtain a plane that is orthogonal to the vector $\mathbf{w}$
- The inequalities $\mathbf{w} \cdot \mathbf{v}>1$ and $\mathbf{w} \cdot \mathbf{v}<1$ again define the two subspaces that are created by the plane

- The same concept applies to high-dimensional spaces as well


## Niatrices

## Matrices

- Matrix is a rectangular array of real-valued scalars arranged in $m$ horizontal rows and $n$ vertical columns
- Each element $a_{i j}$ belongs to the $i^{\text {th }}$ row and $j^{\text {th }}$ column
- The elements are denoted $a_{i j}$ or $\mathbf{A}_{i j}$ or $[\mathbf{A}]_{i j}$ or $\mathbf{A}(\boldsymbol{i}, \boldsymbol{j})$

$$
\mathbf{A}=\left[\begin{array}{cccc}
a_{11} & a_{12} & \cdots & a_{1 n} \\
a_{21} & a_{22} & \cdots & a_{2 n} \\
\vdots & \vdots & \ddots & \vdots \\
a_{m 1} & a_{m 2} & \cdots & a_{m n}
\end{array}\right]
$$

- For the matrix $\mathbf{A} \in \mathbb{R}^{m \times n}$, the size (dimension) is $m \times n$ or $(m, n)$
- Matrices are denoted by bold-font upper-case letters


## Niatrices

## Matrices

- Addition or subtraction $\quad(\mathbf{A} \pm \mathbf{B})_{i, j}=\mathbf{A}_{i, j} \pm \mathbf{B}_{i, j}$

$$
\left[\begin{array}{lll}
1 & 3 & 1 \\
1 & 0 & 0
\end{array}\right]+\left[\begin{array}{lll}
0 & 0 & 5 \\
7 & 5 & 0
\end{array}\right]=\left[\begin{array}{lll}
1+0 & 3+0 & 1+5 \\
1+7 & 0+5 & 0+0
\end{array}\right]=\left[\begin{array}{lll}
1 & 3 & 6 \\
8 & 5 & 0
\end{array}\right]
$$

- Scalar multiplication

$$
(c \mathbf{A})_{i, j}=c \cdot \mathbf{A}_{i, j}
$$

$$
2 \cdot\left[\begin{array}{ccc}
1 & 8 & -3 \\
4 & -2 & 5
\end{array}\right]=\left[\begin{array}{ccc}
2 \cdot 1 & 2 \cdot 8 & 2 \cdot-3 \\
2 \cdot 4 & 2 \cdot-2 & 2 \cdot 5
\end{array}\right]=\left[\begin{array}{ccc}
2 & 16 & -6 \\
8 & -4 & 10
\end{array}\right]
$$

- Matrix multiplication ( $\mathbf{A B})_{i, j}=\mathbf{A}_{i, 1} \mathbf{B}_{1, j}+\mathbf{A}_{i, 2} \mathbf{B}_{2, j}+\cdots+\mathbf{A}_{i, n} \mathbf{B}_{n, j}$
- Defined only if the number of columns of the left matrix is the same as the number of rows of the right matrix
- Note that $\mathbf{A B} \neq \mathbf{B A}$

$$
\left[\begin{array}{ccc}
\underline{2} & \underline{3} & \underline{4} \\
1 & 0 & 0
\end{array}\right]\left[\begin{array}{cc}
0 & \frac{1000}{1} \\
\frac{100}{10} \\
0 & \underline{10}
\end{array}\right]=\left[\begin{array}{ll}
3 & \underline{2340} \\
0 & 1000
\end{array}\right]
$$

## N1atrices

Matrices

- Transpose of the matrix: $\mathbf{A}^{T}$ has the rows and columns exchanged

$$
\left(\mathbf{A}^{T}\right)_{i, j}=\mathbf{A}_{j, i} \quad\left[\begin{array}{ccc}
1 & 2 & 3 \\
0 & -6 & 7
\end{array}\right]^{\mathrm{T}}=\left[\begin{array}{cc}
1 & 0 \\
2 & -6 \\
3 & 7
\end{array}\right]
$$

- Some properties $\mathbf{A}+\mathbf{B}=\mathbf{B}+\mathbf{A}$

$$
\mathbf{A}(\mathbf{B}+\mathbf{C})=\mathbf{A B}+\mathbf{A C}
$$

$$
(\mathbf{A}+\mathbf{B})^{T}=\mathbf{A}^{T}+\mathbf{B}^{T} \quad \mathbf{A}(\mathbf{B C})=(\mathbf{A B}) \mathbf{C}
$$

$$
\left(\mathbf{A}^{T}\right)^{T}=\mathbf{A} \quad(\mathbf{A B})^{T}=\boldsymbol{B}^{T} \mathbf{A}^{T}
$$

- Square matrix: has the same number of rows and columns
- Identity matrix ( $\mathbf{I}_{n}$ ): has ones on the main diagonal, and zeros elsewhere
- E.g.: identity matrix of size $3 \times 3$ : $\mathbf{I}_{3}=\left[\begin{array}{lll}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right]$


## Natrices

## Matrices

- Determinant of a matrix, denoted by $\operatorname{det}(\mathbf{A})$ or $|\mathbf{A}|$, is a real-valued scalar encoding certain properties of the matrix
- E.g., for a matrix of size $2 \times 2$ :

$$
\operatorname{det}\left(\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]\right)=a d-b c
$$

- For larger-size matrices the determinant of a matrix id calculated as

$$
\operatorname{det}(\mathbf{A})=\sum_{j} a_{i j}(-1)^{i+j} \operatorname{det}\left(\mathbf{A}_{(i, j)}\right)
$$

- In the above, $\mathbf{A}_{(i, j)}$ is a minor of the matrix obtained by removing the row and column associated with the indices $i$ and $j$
- Trace of a matrix is the sum of all diagonal elements

$$
\operatorname{Tr}(\mathbf{A})=\sum_{i} a_{i i}
$$

- A matrix for which $\mathbf{A}=\mathbf{A}^{T}$ is called a symmetric matrix


## Matrices

Matrices

- Elementwise multiplication of two matrices A and B is called the Hadamard product or elementwise product
- The math notation is $\odot$

$$
\mathbf{A} \odot \mathbf{B}=\left[\begin{array}{cccc}
a_{11} b_{11} & a_{12} b_{12} & \ldots & a_{1 n} b_{1 n} \\
a_{21} b_{21} & a_{22} b_{22} & \ldots & a_{2 n} b_{2 n} \\
\vdots & \vdots & \ddots & \vdots \\
a_{m 1} b_{m 1} & a_{m 2} b_{m 2} & \ldots & a_{m n} b_{m n}
\end{array}\right]
$$

## Matrix-Vector Products

## Matrices

- Consider a matrix $\mathbf{A} \in \mathbb{R}^{m \times n}$ and a vector $\mathbf{x} \in \mathbb{R}^{n}$
- The matrix can be written in terms of its row vectors (e.g., $\mathbf{a}_{1}^{T}$ is the first row)

$$
\mathbf{A}=\left[\begin{array}{c}
\mathbf{a}_{1}^{\top} \\
\mathbf{a}_{2}^{\top} \\
\vdots \\
\mathbf{a}_{m}^{\top}
\end{array}\right]
$$

- The matrix-vector product is a column vector of length $m$, whose $i^{\text {th }}$ element is the $\operatorname{dot}$ product $\mathbf{a}_{i}^{T} \mathbf{x}$

$$
\mathbf{A} \mathbf{x}=\left[\begin{array}{c}
\mathbf{a}_{1}^{\top} \\
\mathbf{a}_{2}^{\top} \\
\vdots \\
\mathbf{a}_{m}^{\top}
\end{array}\right] \mathbf{x}=\left[\begin{array}{c}
\mathbf{a}_{1}^{\top} \mathbf{x} \\
\mathbf{a}_{2}^{\top} \mathbf{x} \\
\vdots \\
\mathbf{a}_{m}^{\top} \mathbf{x}
\end{array}\right]
$$

- Note the size: $\mathbf{A}(m \times n) \cdot \mathbf{x}(n \times 1)=\mathbf{A x}(m \times 1)$


## Matrix-Matrix Products

## Matrices

- To multiply two matrices $\mathbf{A} \in \mathbb{R}^{n \times k}$ and $\mathbf{B} \in \mathbb{R}^{k \times m}$

$$
\mathbf{A}=\left[\begin{array}{cccc}
a_{11} & a_{12} & \cdots & a_{1 k} \\
a_{21} & a_{22} & \cdots & a_{2 k} \\
\vdots & \vdots & \ddots & \vdots \\
a_{n 1} & a_{n 2} & \cdots & a_{n k}
\end{array}\right], \quad \mathbf{B}=\left[\begin{array}{cccc}
b_{11} & b_{12} & \cdots & b_{1 m} \\
b_{21} & b_{22} & \cdots & b_{2 m} \\
\vdots & \vdots & \ddots & \vdots \\
b_{k 1} & b_{k 2} & \cdots & b_{k m}
\end{array}\right]
$$

- We can consider the matrix-matrix product as dot-products of rows in $\mathbf{A}$ and columns in B

$$
\mathbf{C}=\mathbf{A B}=\left[\begin{array}{c}
\mathbf{a}_{1}^{\top} \\
\mathbf{a}_{2}^{\top} \\
\vdots \\
\mathbf{a}_{n}^{\top}
\end{array}\right]\left[\begin{array}{llll}
\mathbf{b}_{1} & \mathbf{b}_{2} & \cdots & \mathbf{b}_{m}
\end{array}\right]=\left[\begin{array}{cccc}
\mathbf{a}_{1}^{\top} \mathbf{b}_{1} & \mathbf{a}_{1}^{\top} \mathbf{b}_{2} & \cdots & \mathbf{a}_{1}^{\top} \mathbf{b}_{m} \\
\mathbf{a}_{2}^{\top} \mathbf{b}_{1} & \mathbf{a}_{2}^{\top} \mathbf{b}_{2} & \cdots & \mathbf{a}_{2}^{\top} \mathbf{b}_{m} \\
\vdots & \vdots & \ddots & \vdots \\
\mathbf{a}_{n}^{\top} \mathbf{b}_{1} & \mathbf{a}_{n}^{\top} \mathbf{b}_{2} & \cdots & \mathbf{a}_{n}^{\top} \mathbf{b}_{m}
\end{array}\right]
$$

- Size: $\mathbf{A}(n \times k) \cdot \mathbf{B}(k \times m)=\mathbf{C}(n \times m)$


## Inverse of a Matrix

## Matrices

- For a square $n \times n$ matrix $\mathbf{A}$ with rank $n, \mathbf{A}^{\mathbf{1}}$ is its inverse matrix if their product is an identity matrix I

$$
\mathbf{A}^{-1} \mathbf{A}=\mathbf{A} \mathbf{A}^{-1}=\mathbf{I}
$$

- Properties of inverse matrices $\left(\mathbf{A}^{-1}\right)^{-1}=\mathbf{A}$

$$
(\mathbf{A B})^{-1}=\mathbf{B}^{-1} \mathbf{A}^{-1}
$$

- If $\operatorname{det}(A)=0$ (i.e., $\operatorname{rank}(A)<n$ ), then the inverse does not exist
- A matrix that is not invertible is called a singular matrix
- Note that finding an inverse of a large matrix is computationally expensive
- In addition, it can lead to numerical instability
- If the inverse of a matrix is equal to its transpose, the matrix is said to be orthogonal matrix

$$
\mathbf{A}^{-1}=\mathbf{A}^{T}
$$

## Pseudo-Inverse of a Matrix

## Matrices

- Pseudo-inverse of a matrix
- Also known as Moore-Penrose pseudo-inverse
- For matrices that are not square, the inverse does not exist
- Therefore, a pseudo-inverse is used
- If $m>n$, then the pseudo-inverse is $\mathbf{A}^{\dagger}=\left(\mathbf{A}^{T} \mathbf{A}\right)^{-1} \mathbf{A}^{T}$ and $\mathbf{A}^{\dagger} \mathbf{A}=\mathbf{I}$
- If $m<n$, then the pseudo-inverse is $\mathbf{A}^{\dagger}=\mathbf{A}^{T}\left(\mathbf{A} \mathbf{A}^{T}\right)^{-1}$ and $\mathbf{A} \mathbf{A}^{\dagger}=\mathbf{I}$
- E.g., for a matrix with dimension $\mathbf{X}_{2 \times 3}$, a pseudo-inverse can be found of size $\mathbf{X}_{3 \times 2}^{\dagger}$, so that $\mathbf{X}_{2 \times 3} \mathbf{X}_{3 \times 2}^{\dagger}=\mathbf{I}_{2 \times 2}$


## Tensors

- Tensors are $n$-dimensional arrays of scalars
- Vectors are first-order tensors, $\mathbf{v} \in \mathbb{R}^{n}$
- Matrices are second-order tensors, $\mathbf{A} \in \mathbb{R}^{m \times n}$
- E.g., a fourth-order tensor is $\mathbf{T} \in \mathbb{R}^{n_{1} \times n_{2} \times n_{3} \times n_{4}}$
- Tensors are denoted with upper-case letters of a special font face (e.g., X, Y, Z )
- RGB images are third-order tensors, i.e., as they are 3-dimensional arrays
- The 3 axes correspond to width, height, and channel
- E.g., $224 \times 224 \times 3$
- The channel axis corresponds to the color channels (red, green, and blue)


## Manifolds

## Manifolds

- Earlier we learned that hyperplanes generalize the concept of planes in highdimensional spaces
- Similarly, manifolds can be informally imagined as generalization of the concept of surfaces in high-dimensional spaces
- To begin with an intuitive explanation, the surface of the Earth is an example of a two-dimensional manifold embedded in a three-dimensional space
- This is true because the Earth looks locally flat, so on a small scale it is like a 2-D plane
- However, if we keep walking on the Earth in one direction, we will eventually end up back where we started
- This means that Earth is not really flat, it only looks locally like a Euclidean plane, but at large scales it folds up on itself, and has a different global structure than a flat plane


## Manifolds

## Manifolds

- Manifolds are studied in mathematics under topological spaces
- An $n$-dimensional manifold is defined as a topological space with the property that each point has a neighborhood that is homeomorphic to the Euclidean space of dimension $n$
- This means that a manifold locally resembles Euclidean space near each point
- Informally, a Euclidean space is locally smooth, it does not have holes, edges, or other sudden changes, and it does not have intersecting neighborhoods
- Although the manifolds can have very complex structure on a large scale, resemblance of the Euclidean space on a small scale allows to apply standard math concepts
- Examples of 2-dimensional manifolds are shown in the figure
- The surfaces in the figure have been conveniently cut up into little rectangles that were glued together
- Those small rectangles locally look like flat

Sphere


Double torus

Torus


Klein bottle Euclidean planes

## Manifolds

## Manifolds

- Examples of one-dimensional manifolds
- Upper figure: a circle is a l-D manifold embedded in 2-D, where each arc of the circle locally resembles a line segment
- Lower figures: other examples of 1-D manifolds
- Note that a number 8 figure is not a manifold because it has an intersecting point (it is not Euclidean locally)
- It is hypothesized that in the real-world, high-dimensional
 data (such as images) lie on low-dimensional manifolds embedded in the high-dimensional space
- E.g., in ML, let's assume we have a training set of images with size $224 \times 224 \times 3$ pixels
- Learning an arbitrary function in such high-dimensional space would be intractable

" Despite that, all images of the same class ("cats" for example) might lie on a low-dimensional manifold
- This allows function learning and image classification


## Manifolds

## Manifolds

- Example:
- The data points have 3 dimensions (left figure), i.e., the input space of the data is 3dimensional
- The data points lie on a 2-dimensional manifold, shown in the right figure
- Most ML algorithms extract lower-dimensional data features that enable to distinguish between various classes of high-dimensional input data
- The low-dimensional representations of the input data are called embeddings




## Eigen Decomposition

Eigen Decomposition

- Eigen decomposition is decomposing a matrix into a set of eigenvalues and eigenvectors
- Eigenvalues of a square matrix $\mathbf{A}$ are scalars $\lambda$ and eigenvectors are non-zero vectors $\mathbf{v}$ that satisfy

$$
\mathbf{A} \mathbf{v}=\lambda \mathbf{v}
$$

- Eigenvalues are found by solving the following equation

$$
\operatorname{det}(\mathbf{A}-\lambda \mathbf{I})=0
$$

- If a matrix $\mathbf{A}$ has $n$ linearly independent eigenvectors $\left\{\mathbf{v}^{1}, \ldots, \mathbf{v}^{n}\right\}$ with corresponding eigenvalues $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$, the eigen decomposition of $\mathbf{A}$ is given by

$$
\mathbf{A}=\mathbf{V} \boldsymbol{\Lambda} \mathbf{V}^{-1}
$$

- Columns of the matrix $\mathbf{V}$ are the eigenvectors, i.e., $\mathbf{V}=\left[\mathbf{v}^{1}, \ldots, \mathbf{v}^{n}\right]$
- $\boldsymbol{\Lambda}$ is a diagonal matrix of the eigenvalues, i.e., $\boldsymbol{\Lambda}=\left[\lambda_{1}, \ldots, \lambda_{n}\right]$
- To find the inverse of the matrix $A$, we can use $\mathbf{A}^{\mathbf{- 1}}=\mathbf{V} \boldsymbol{\Lambda}^{-\mathbf{1}} \mathbf{V}^{-1}$
- This involves simply finding the inverse $\boldsymbol{\Lambda}^{\mathbf{- 1}}$ of a diagonal matrix


## Eigen Decomposition

Eigen Decomposition

- Decomposing a matrix into eigenvalues and eigenvectors allows to analyze certain properties of the matrix
- If all eigenvalues are positive, the matrix is positive definite
- If all eigenvalues are positive or zero-valued, the matrix is positive semidefinite
- If all eigenvalues are negative or zero-values, the matrix is negative semidefinite
- Positive semidefinite matrices are interesting because they guarantee that $\forall \mathbf{x}, \mathbf{x}^{T} \mathbf{A x} \geq 0$
- Eigen decomposition can also simplify many linear-algebraic computations
- The determinant of A can be calculated as

$$
\operatorname{det}(\mathbf{A})=\lambda_{1} \cdot \lambda_{2} \cdots \lambda_{n}
$$

- If any of the eigenvalues are zero, the matrix is singular (it does not have an inverse)
- However, not every matrix can be decomposed into eigenvalues and eigenvectors
- Also, in some cases the decomposition may involve complex numbers
- Still, every real symmetric matrix is guaranteed to have an eigen decomposition according to $\mathbf{A}=\mathbf{V} \boldsymbol{\Lambda} \mathbf{V}^{-1}$, where $\mathbf{V}$ is an orthogonal matrix


## Eigen Decomposition

Eigen Decomposition

- Geometric interpretation of the eigenvalues and eigenvectors is that they allow to stretch the space in specific directions
- Left figure: the two eigenvectors $\mathbf{v}^{1}$ and $\mathbf{v}^{2}$ are shown for a matrix, where the two vectors are unit vectors (i.e., they have a length of 1)
- Right figure: the vectors $\mathbf{v}^{1}$ and $\mathbf{v}^{2}$ are multiplied with the eigenvalues $\lambda_{1}$ and $\lambda_{2}$
- We can see how the space is scaled in the direction of the larger eigenvalue $\lambda_{1}$
- E.g., this is used for dimensionality reduction with PCA (principal component analysis) where the eigenvectors corresponding to the largest eigenvalues are used for extracting the most important data dimensions



## Differential Calculus

## Differential Calculus

- For a function $f: \mathbb{R} \rightarrow \mathbb{R}$, the derivative of $f$ is defined as

$$
f^{\prime}(x)=\lim _{h \rightarrow 0} \frac{f(x+h)-f(x)}{h}
$$

- If $f^{\prime}(a)$ exists, $f$ is said to be differentiable at $a$
- If $f^{\prime}(c)$ is differentiable for $\forall c \in[a, b]$, then $f$ is differentiable on this interval
- We can also interpret the derivative $f^{\prime}(x)$ as the instantaneous rate of change of $f(x)$ with respect to $x$
- I.e., for a small change in $x$, what is the rate of change of $f(x)$
- Given $y=f(x)$, where $x$ is an independent variable and $y$ is a dependent variable, the following expressions are equivalent:

$$
f^{\prime}(x)=f^{\prime}=\frac{d y}{d x}=\frac{d f}{d x}=\frac{d}{d x} f(x)=D f(x)=D_{x} f(x)
$$

- The symbols $\frac{d}{d x^{\prime}} D$, and $D_{x}$ are differentiation operators that indicate operation of differentiation


## Differential Calculus

Differential Calculus

- The following rules are used for computing the derivatives of explicit functions
- Derivative of constants. $\frac{d}{d x} c=0$.
- Derivative of linear functions. $\frac{d}{d x}(a x)=a$.
- Power rule. $\frac{d}{d x} x^{n}=n x^{n-1}$.
- Derivative of exponentials. $\frac{d}{d x} e^{x}=e^{x}$.
- Derivative of the logarithm. $\frac{d}{d x} \log (x)=\frac{1}{x}$.
- Sum rule. $\frac{d}{d x}(g(x)+h(x))=\frac{d g}{d x}(x)+\frac{d h}{d x}(x)$.
- Product rule. $\frac{d}{d x}(g(x) \cdot h(x))=g(x) \frac{d h}{d x}(x)+\frac{d g}{d x}(x) h(x)$.
- Chain rule. $\frac{d}{d x} g(h(x))=\frac{d g}{d h}(h(x)) \cdot \frac{d h}{d x}(x)$.


## rionerserderserivatives

## Differential Calculus

- The derivative of the first derivative of a function $f(x)$ is the second derivative of $f(x)$

$$
\frac{d^{2} f}{d x^{2}}=\frac{d}{d x}\left(\frac{d f}{d x}\right)
$$

- The second derivative quantifies how the rate of change of $f(x)$ is changing
- E.g., in physics, if the function describes the displacement of an object, the first derivative gives the velocity of the object (i.e., the rate of change of the position)
- The second derivative gives the acceleration of the object (i.e., the rate of change of the velocity)
- If we apply the differentiation operation any number of times, we obtain the $n$-th derivative of $f(x)$

$$
f^{(n)}(x)=\frac{d^{n} f}{d x^{n}}=\left(\frac{d}{d x}\right)^{n} f(x)
$$

## Geonetric intericetation

## Differential Calculus

- To provide a geometric interpretation of the derivatives, let's consider a firstorder Taylor series approximation of $f(x)$ at $x=x_{0}$

$$
f(x) \approx f\left(x_{0}\right)+\left.\frac{d f}{d x}\right|_{x_{0}}\left(x-x_{0}\right)
$$

- The expression approximates the function $f(x)$ by a line which passes through the point $\left(x_{0}, f\left(x_{0}\right)\right)$ and has slope $\left.\frac{d f}{d x}\right|_{x_{0}}$ (i.e., the value of $\frac{d f}{d x}$ at the point $x_{0}$ )
- Therefore, the first derivative of a function is also the slope of the tangent line to the curve of the function



## Partial Derivatives

## Differential Calculus

- So far, we looked at functions of a single variable, where $f: \mathbb{R} \rightarrow \mathbb{R}$
- Functions that depend on many variables are called multivariate functions
- Let $y=f(\mathbf{x})=f\left(x_{1}, x_{2}, \ldots, x_{n}\right)$ be a multivariate function with $n$ variables
- The input is an $n$-dimensional vector $\mathbf{x}=\left[x_{1}, x_{2}, \ldots, x_{n}\right]^{T}$ and the output is a scalar $y$
- The mapping is $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$
- The partial derivative of $y$ with respect to its $i^{\text {th }}$ parameter $x_{i}$ is

$$
\frac{\partial y}{\partial x_{i}}=\lim _{h \rightarrow 0} \frac{f\left(x_{1}, x_{2}, \ldots, x_{i}+h, \ldots, x_{n}\right)-f\left(x_{1}, x_{2}, \ldots, x_{i}, \ldots, x_{n}\right)}{h}
$$

- To calculate $\frac{\partial y}{\partial x_{i}}(\partial$ pronounced "del" or we can just say "partial derivative"), we can treat $x_{1}, x_{2}, \ldots, x_{i-1}, \quad x_{i+1} \ldots, x_{n}$ as constants and calculate the derivative of $y$ only with respect to $x_{i}$
- For notation of partial derivatives, the following are equivalent:

$$
\frac{\partial y}{\partial x_{i}}=\frac{\partial f}{\partial x_{i}}=\frac{\partial}{\partial x_{i}} f(\mathbf{x})=f_{x_{i}}=f_{i}=D_{i} f=D_{x_{i}} f
$$

## Gradient

## Differential Calculus

- We can concatenate partial derivatives of a multivariate function with respect to all its input variables to obtain the gradient vector of the function
- The gradient of the multivariate function $f(\mathbf{x})$ with respect to the $n$-dimensional input vector $\mathbf{x}=\left[x_{1}, x_{2}, \ldots, x_{n}\right]^{T}$, is a vector of $n$ partial derivatives

$$
\nabla_{\mathbf{x}} f(\mathbf{x})=\left[\frac{\partial f(\mathbf{x})}{\partial x_{1}}, \frac{\partial f(\mathbf{x})}{\partial x_{2}}, \ldots, \frac{\partial f(\mathbf{x})}{\partial x_{n}}\right]^{T}
$$

- When there is no ambiguity, the notations $\nabla f(\mathbf{x})$ or $\nabla_{\mathbf{x}} f$ are often used for the gradient instead of $\nabla_{\mathbf{x}} f(\mathbf{x})$
" The symbol for the gradient is the Greek letter $\nabla$ (pronounced "nabla"), although $\nabla_{\mathbf{x}} f(\mathbf{x})$ is more often it is pronounced "gradient of $f$ with respect to $\mathbf{x}$ "
- In ML, the gradient descent algorithm relies on the opposite direction of the gradient of the loss function $\mathcal{L}$ with respect to the model parameters $\theta\left(\nabla_{\theta} \mathcal{L}\right)$ for minimizing the loss function
- Adversarial examples can be created by adding perturbation in the direction of the gradient of the loss $\mathcal{L}$ with respect to input examples $x\left(\nabla_{x} \mathcal{L}\right)$ for maximizing the loss function


## Optimization

Optimization

- Optimization is concerned with optimizing an objective function - finding the value of an argument that minimizes or maximizes the function
- Most optimization algorithms are formulated in terms of minimizing a function $f(x)$
- Maximization is accomplished vie minimizing the negative of an objective function (e.g., minimize $-f(x)$ )
- In minimization problems, the objective function is often referred to as a cost function or loss function or error function
- Optimization is very important for machine learning
- The performance of optimization algorithms affect the model's training efficiency
- Most optimization problems in machine learning are nonconvex
- Meaning that the loss function is not a convex function
- Nonetheless, the design and analysis of algorithms for solving convex problems has been very instructive for advancing the field of machine learning


## Optimization

Optimization

- Optimization and machine learning have related, but somewhat different goals
- Goal in optimization: minimize an objective function
- For a set of training examples, reduce the training error
- Goal in ML: find a suitable model, to predict on data examples
- For a set of testing examples, reduce the generalization error
- For a given empirical function $g$ (dashed purple curve), optimization algorithms attempt to find the point of minimum empirical risk
- The expected function $f$ (blue curve) is obtained given a limited amount of training data examples
- ML algorithms attempt to find the point of minimum expected risk, based on minimizing the error on a set of testing examples
- Which may be at a different location than the minimum of the training examples
- And which may not be minimal in a formal sense



## Stationary Points

## Optimization

- Stationary points ( or critical points) of a differentiable function $f(x)$ of one variable are the points where the derivative of the function is zero, i.e., $f^{\prime}(x)=0$
- The stationary points can be:
- Minimum, a point where the derivative changes from negative to positive
- Maximum, a point where the derivative changes from positive to negative
- Saddle point, derivative is either positive or negative on both sides of the point
- The minimum and maximum points are collectively known as extremum points
- The nature of stationary points can be determined based on the second derivative of $f(x)$ at the point
- If $f^{\prime \prime}(x)>0$, the point is a minimum
- If $f^{\prime \prime}(x)<0$, the point is a maximum
- If $f^{\prime \prime}(x)=0$, inconclusive, the point can be a saddle point, but it may not
- The same concept also applies to gradients of multivariate functions



## Local Minima

## Optimization

- Among the challenges in optimization of model's parameters in ML involve local minima, saddle points, vanishing gradients
- For an objective function $f(x)$, if the value at a point $x$ is the minimum of the objective function over the entire domain of $x$, then it is the global minimum
- If the value of $f(x)$ at $x$ is smaller than the values of the objective function at any other points in the vicinity of $x$, then it is the local minimum
- The objective functions in ML usually have many local minima
- When the solution of the optimization algorithm is near the local minimum, the gradient of the loss function approaches or becomes zero (vanishing gradients)
- Therefore, the obtained solution in the final iteration can be a local minimum, rather than the global minimum



## Saddle Points

## Optimization

- The gradient of a function $f(x)$ at a saddle point is 0 , but the point is not a minimum or maximum point
- The optimization algorithms may stall at saddle points, without reaching a minima
- Note also that the point of a function at which the sign of the curvature changes is called an inflection point
- An inflection point $\left(f^{\prime \prime}(x)=0\right)$ can also be a saddle point, but it does not have to be
- For the 2D function (right figure), the saddle point is at $(0,0)$
- The point looks like a saddle, and gives the minimum with respect to $x$, and the maximum with respect to $y$




## Convex Optimization

Optimization

- A function of a single variable is concave if every line segment joining two points on its graph does not lie above the graph at any point
- Symmetrically, a function of a single variable is convex if every line segment joining two points on its graph does not lie below the graph at any point


A concave function: no line segment joining two points on the graph lies above the graph at any point


A convex function: no line segment joining two points on the graph lies below the graph at any point


A function that is neither concave nor convex: the line segment shown lies above the graph at some points and below it at others

## Convex Functions

Optimization

- In mathematical terms, the function $f$ is a convex function if for all points $x_{1}, x_{2}$ and for all $\lambda \in[0,1]$

$$
\lambda f\left(x_{1}\right)+(1-\lambda) f\left(x_{2}\right) \geq f\left(\lambda x_{1}+(1-\lambda) x_{2}\right)
$$



## Convex Functions

Optimization

- One important property of convex functions is that they do not have local minima
- Every local minimum of a convex function is a global minimum
- I.e., every point at which the gradient of a convex function $=0$ is the global minimum
- The figure below illustrates two convex functions, and one nonconvex function



## Probability

## Probability

- Intuition:
- In a process, several outcomes are possible
- When the process is repeated a large number of times, each outcome occurs with a relative frequency, or probability
- If a particular outcome occurs more often, we say it is more probable
- Probability arises in two contexts
- In actual repeated experiments
- Example: You record the color of 1,000 cars driving by. 57 of them are green. You estimate the probability of a car being green as $57 / 1,000=0.057$.
- In idealized conceptions of a repeated process
- Example: You consider the behavior of an unbiased six-sided die. The expected probability of rolling a 5 is $1 / 6=0.1667$.
- Example: You need a model for how people's heights are distributed. You choose a normal distribution to represent the expected relative probabilities.


## Probability

## Probability

- Solving machine learning problems requires to deal with uncertain quantities, as well as with stochastic (non-deterministic) quantities
- Probability theory provides a mathematical framework for representing and quantifying uncertain quantities
- There are different sources of uncertainty:
- Inherent stochasticity in the system being modeled
- For example, most interpretations of quantum mechanics describe the dynamics of subatomic particles as being probabilistic
- Incomplete observability
- Even deterministic systems can appear stochastic when we cannot observe all of the variables that drive the behavior of the system
- Incomplete modeling
- When we use a model that must discard some of the information we have observed, the discarded information results in uncertainty in the model's predictions
- E.g., discretization of real-numbered values, dimensionality reduction, etc.


## Random variables

Probability

- A random variable $X$ is a variable that can take on different values
- Example: $X=$ rolling a die
- Possible values of $X$ comprise the sample space, or outcome space, $\mathcal{S}=\{1,2,3,4,5,6\}$
- We denote the event of "seeing a 5" as $\{X=5\}$ or $X=5$
- The probability of the event is $P(\{X=5\})$ or $P(X=5)$
- Also, $P(5)$ can be used to denote the probability that $X$ takes the value of 5
- A probability distribution is a description of how likely a random variable is to take on each of its possible states
- A compact notation is common, where $P(X)$ is the probability distribution over the random variable $X$
- Also, the notation $\mathrm{X} \sim P(X)$ can be used to denote that the random variable $X$ has probability distribution $P(X)$
- Random variables can be discrete or continuous
- Discrete random variables have finite number of states: e.g., the sides of a die
- Continuous random variables have infinite number of states: e.g., the height of a person


## Axioms of probability

Probability

- The probability of an event $\mathcal{A}$ in the given sample space $\mathcal{S}$, denoted as $P(\mathcal{A})$, must satisfies the following properties:
- Non-negativity
- For any event $\mathcal{A} \in \mathcal{S}, P(\mathcal{A}) \geq 0$
- All possible outcomes
- Probability of the entire sample space is $1, P(S)=1$
- Additivity of disjoint events
$\circ$ For all events $\mathcal{A}_{1}, \mathcal{A}_{2} \in \mathcal{S}$ that are mutually exclusive $\left(\mathcal{A}_{1} \cap \mathcal{A}_{2}=\varnothing\right.$ ), the probability that both events happen is equal to the sum of their individual probabilities, $P\left(\mathcal{A}_{1} \cup \mathcal{A}_{2}\right)=$ $P\left(\mathcal{A}_{1}\right)+P\left(\mathcal{A}_{2}\right)$
- The probability of a random variable $P(X)$ must obey the axioms of probability over the possible values in the sample space $\mathcal{S}$


## Discrete Variables

Probability

- A probability distribution over discrete variables may be described using a probability mass function (PMF)
- E.g., sum of two dice
- A probability distribution over continuous
 variables may be described using a probability density function (PDF)
- E.g., waiting time between eruptions of Old Faithful
- A PDF gives the probability of an infinitesimal region with volume $\delta X$
- To find the probability over an interval [ $a, b$ ], we can integrate the PDF as follows:


$$
P(X \in[a, b])=\int_{a}^{b} P(X) d X
$$

## Niultivariate Randon Variabies

## Probability

- We may need to consider several random variables at a time
- If several random processes occur in parallel or in sequence
- E.g., to model the relationship between several diseases and symptoms
- E.g., to process images with millions of pixels (each pixel is one random variable)
- Next, we will study probability distributions defined over multiple random variables
- These include joint, conditional, and marginal probability distributions
- The individual random variables can also be grouped together into a random vector, because they represent different properties of an individual statistical unit
- A multivariate random variable is a vector of multiple random variables $\mathbf{X}=$ $\left(X_{1}, X_{2}, \ldots, X_{n}\right)^{T}$


## Joint Probability Distribution

## Probability

- Probability distribution that acts on many variables at the same time is known as a joint probability distribution
- Given any values $x$ and $y$ of two random variables $X$ and $Y$, what is the probability that $X=x$ and $Y=y$ simultaneously?
- $P(X=x, Y=y)$ denotes the joint probability
- We may also write $P(x, y)$ for brevity



## Marginal Probability Distribution

Probability

- Marginal probability distribution is the probability distribution of a single variable
- It is calculated based on the joint probability distribution $P(X, Y)$
- I.e., using the sum rule: $P(X=x)=\sum_{y} P(X=x, Y=y)$
- For continuous random variables, the summation is replaced with integration, $P(X=x)=$ $\int P(X=x, Y=y) d y$
- This process is called marginalization



## Conditional Probability Distribution

Probability

- Conditional probability distribution is the probability distribution of one variable provided that another variable has taken a certain value
- Denoted $P(X=x \mid Y=y)$
- Note that: $P(X=x \mid Y=y)=\frac{P(X=x, Y=y)}{P(Y=y)}$
conditional probability: $p(Y=$ European $\mid X=$ minivan $)=$
$0.1481 /(0.0741+0.1111+0.1481)=0.4433$



## Bayes'rneorenn

Probability

- Bayes' theorem - allows to calculate conditional probabilities for one variable when conditional probabilities for another variable are known

$$
P(X \mid Y)=\frac{P(Y \mid X) P(X)}{P(Y)}
$$

- Also known as Bayes' rule
- Multiplication rule for the joint distribution is used: $P(X, Y)=P(Y \mid X) P(X)$
- By symmetry, we also have: $P(Y, X)=P(X \mid Y) P(Y)$
- The terms are referred to as:
- $P(X)$, the prior probability, the initial degree of belief for $X$
- $P(X \mid Y)$, the posterior probability, the degree of belief after incorporating the knowledge of $Y$
- $P(Y \mid X)$, the likelihood of $Y$ given $X$
- $\mathrm{P}(\mathrm{Y})$, the evidence
- Bayes' theorem: posterior probability $=\frac{\text { likelihood } \times \text { prior probability }}{\text { evidence }}$


## Independence

Probability

- Two random variables $X$ and $Y$ are independent if the occurrence of $Y$ does not reveal any information about the occurrence of $X$
- E.g., two successive rolls of a die are independent
- Therefore, we can write: $P(X \mid Y)=P(X)$
- The following notation is used: $X \perp Y$
- Also note that for independent random variables: $P(X, Y)=P(X) P(Y)$
- In all other cases, the random variables are dependent
- E.g., duration of successive eruptions of Old Faithful
- Getting a king on successive draws form a deck (the drawn card is not replaced)
- Two random variables $X$ and $Y$ are conditionally independent given another random variable $Z$ if and only if $P(X, Y \mid Z)=P(X \mid Z) P(Y \mid Z)$
- This is denoted as $X \perp Y \mid Z$


## Expected Value

Probability

- The expected value or expectation of a function $f(X)$ with respect to a probability distribution $P(X)$ is the average (mean) when $X$ is drawn from $P(X)$
- For a discrete random variable $X$, it is calculated as

$$
\mathbb{E}_{X \sim P}[f(X)]=\sum_{X} P(X) f(X)
$$

- For a continuous random variable $X$, it is calculated as

$$
\mathbb{E}_{X \sim P}[f(X)]=\int P(X) f(X) d X
$$

- When the identity of the distribution is clear from the context, we can write $\mathbb{E}_{X}[f(X)]$
- If it is clear which random variable is used, we can write just $\mathbb{E}[f(X)]$
- Mean is the most common measure of central tendency of a distribution
- For a random variable: $f\left(X_{i}\right)=X_{i} \quad \Rightarrow \quad \mu=\mathbb{E}\left[X_{i}\right]=\sum_{i} P\left(X_{i}\right) \cdot X_{i}$
- This is similar to the mean of a sample of observations: $\mu=\frac{1}{N} \sum_{i} X_{i}$
- Other measures of central tendency: median, mode


## Variance

Probability

- Variance gives the measure of how much the values of the function $f(X)$ deviate from the expected value as we sample values of $X$ from $P(X)$

$$
\operatorname{Var}(f(X))=\mathbb{E}\left[(f(X)-\mathbb{E}[f(X)])^{2}\right]
$$

- When the variance is low, the values of $f(X)$ cluster near the expected value
- Variance is commonly denoted with $\sigma^{2}$
- The above equation is similar to a function $f\left(X_{i}\right)=X_{i}-\mu$
- We have $\sigma^{2}=\sum_{i} P\left(X_{i}\right) \cdot\left(X_{i}-\mu\right)^{2}$
- This is similar to the formula for calculating the variance of a sample of observations: $\sigma^{2}=\frac{1}{N-1} \sum_{i}\left(X_{i}-\mu\right)^{2}$
- The square root of the variance is the standard deviation
- Denoted $\sigma=\sqrt{\operatorname{Var}(X)}$


## Covariance

## Probability

- Covariance gives the measure of how much two random variables are linearly related to each other

$$
\operatorname{Cov}(f(X), g(Y))=\mathbb{E}[(f(X)-\mathbb{E}[f(X)])(g(Y)-\mathbb{E}[g(Y)])]
$$

- If $f\left(X_{i}\right)=X_{i}-\mu_{X}$ and $g\left(Y_{i}\right)=Y_{i}-\mu_{Y}$
- Then, the covariance is: $\operatorname{Cov}(X, Y)=\sum_{i} P\left(X_{i}, Y_{i}\right) \cdot\left(X_{i}-\mu_{X}\right) \cdot\left(Y_{i}-\mu_{Y}\right)$
- Compare to covariance of actual samples: $\operatorname{Cov}(X, Y)=\frac{1}{N-1} \sum_{i}\left(Y_{i}-\mu_{X}\right)\left(Y_{i}-\mu_{Y}\right)$
- The covariance measures the tendency for $X$ and $Y$ to deviate from their means in same (or opposite) directions at same time





## Correlation

Probability

- Correlation coefficient is the covariance normalized by the standard deviations of the two variables

$$
\operatorname{corr}(X, Y)=\frac{\operatorname{Cov}(X, Y)}{\sigma_{X} \cdot \sigma_{Y}}
$$

- It is also called Pearson's correlation coefficient and it is denoted $\rho(X, Y)$
- The values are in the interval $[-1,1]$
- It only reflects linear dependence between variables, and it does not measure nonlinear dependencies between the variables
(2)


## Probability Distributions

- Bernoulli distribution
- Binary random variable $X$ with states $\{0,1\}$
- The random variable can encodes a coin flip which comes up 1 with probability $p$ and 0 with probability $1-p$
- Notation: $X \sim \operatorname{Bernoulli}(p)$

- Uniform distribution
- The probability of each value $i \in\{1,2, \ldots, n\}$ is $p_{i}=\frac{1}{n}$
- Notation: $X \sim U(n)$
- Figure: $n=5, p=0.2$



## Probability Distributions

Probability

- Binomial distribution
- Performing a sequence of $n$ independent experiments, each of which has probability $p$ of succeeding, where $p \in\{0,1\}$
- The probability of getting $k$ successes in $n$ trials is $P(X=k)=\binom{n}{k} p^{k}(1-p)^{n-k}$
- Notation: $X \sim \operatorname{Binomial}(n, p)$
- Poisson distribution
- A number of events occurring independently in a fixed interval of time with a known rate $\lambda$
- A discrete random variable $X$ with states $k \in$ $\{0,1,2, \ldots\}$ has probability $P(X=k)=\frac{\lambda^{X} \cdot e^{-\lambda}}{X!}$
- The rate $\lambda$ is the average number of occurrences of the event
- Notation: $X \sim \operatorname{Poisson}(\lambda)$


## Probability Distributions

## Probability

- Gaussian distribution
- The most well-studied distribution
- Referred to as normal distribution or informally bell-shaped distribution
- Defined with the mean $\mu$ and variance $\sigma^{2}$
- Notation: $X \sim \mathcal{N}\left(\mu, \sigma^{2}\right)$
- For a random variable $X$ with $n$ independent measurements, the density is

$$
P_{X}(x)=\frac{1}{\sqrt{2 \pi \sigma^{2}}} e^{-\frac{(x-\mu)^{2}}{2 \sigma^{2}}}
$$




## Cross-entropy

Information Theory

- Cross-entropy is closely related to the KL divergence, and it is defined as the summation of the entropy $H(P)$ and KL divergence $D_{K L}(P \| Q)$

$$
C E(P, Q)=H(P)+D_{K L}(P \| Q)
$$

- Alternatively, the cross-entropy can be written as

$$
C E(P, Q)=-\mathbb{E}_{X \sim P}[\log Q(X)]
$$

- In machine learning, let's assume a classification problem based on a set of data examples $\left\{x_{1}, x_{2}, \ldots, x_{n}\right\}$, that need to be classified into $k$ classes
- For each data example $x_{i}$ we have a class label $y_{i}$
- The true labels $\mathbf{y}$ follow the true distribution $P$
- The goal is to train a classifier (e.g., a NN) parameterized by $\theta$, that outputs a predicted class label $\hat{y}_{i}$ for each data example $x_{i}$
- The predicted labels $\widehat{\boldsymbol{y}}$ follow the estimated distribution $Q$
- The cross-entropy loss between the true distribution $P$ and the estimated distribution $Q$ is calculated as: $C E(\mathbf{y}, \hat{\mathbf{y}})=-\mathbb{E}_{X \sim P}[\log Q(X)]=-\sum_{X} P(X) \log Q(X)=-\sum_{i} y_{i} \log \hat{y}_{i}$
$\circ$ The further away the true and estimated distributions are, the greater the cross-entropy loss is


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