# Math 516: Numerical Optimization 

Lecture based on<br>Convex Analysis and Nonsmooth Optimization by Dmitriy Drusvyatskiy

## Background Material

## Inner Products

Throughout, $\mathbf{E}$ is a Euclidean space,
i.e., a finite-dim real vector space with an inner product $\langle\cdot, \cdot\rangle$. Occasionally we say that $(\mathbf{E},\langle\cdot, \cdot\rangle)$ is the Euclidean space when the choice of inner product needs to be specified.

Recall that an inner-product on $\mathbf{E}$ is an assignment $\langle\cdot, \cdot\rangle: \mathbf{E} \times \mathbf{E} \rightarrow \mathbf{R}$ satisfying the following three properties for all $x, y, z \in \mathbf{E}$ and scalars $a, b \in \mathbf{R}$ :
(Symmetry) $\langle x, y\rangle=\langle y, x\rangle$
(Bilinearity) $\langle a x+b y, z\rangle=a\langle x, z\rangle+b\langle y, z\rangle$
(Positive definiteness) $\langle x, x\rangle \geq 0$ and equality $\langle x, x\rangle=0$ holds if and only if $x=0$.

## Examples of Inner Products

Standard ip for $\mathbf{R}^{n}:\langle x, y\rangle:=\sum_{i=1}^{n} x_{i} y_{i}=\|x\|\|y\| \cos \theta$, where $\theta$ is the angle between $x$ and $y$.

Standard ip for $\mathbf{R}^{m \times n}$ : The Frobenius or trace inner product,

$$
\langle X, Y\rangle:=\operatorname{tr} X^{T} Y=\sum_{i, j} X_{i j} Y_{i j} .
$$

Real polynomials in one variable of degree $\leq n$ on $[a, b]$ : Integration inner product

$$
\langle p, q\rangle:=\int_{a}^{b} p(t) q(t) d t
$$

## Adjoints of Linear Transformations

Suppose both $\left(\mathbf{X},\langle\cdot, \cdot\rangle_{\mathbf{x}}\right)$ and $\left(\mathbf{Y},\langle\cdot, \cdot\rangle_{\mathbf{Y}}\right)$ are Euclidean spaces.
Let $\mathcal{A} \in \mathbf{L}(\mathbf{X}, \mathbf{Y})$ where $\mathbf{L}(\mathbf{X}, \mathbf{Y})$ is the vector space of linear operators (or linear transformations) from $\mathbf{X}$ to $\mathbf{Y}$.

There exists a unique linear mapping $\mathcal{A}^{*}: \mathbf{Y} \rightarrow \mathbf{X}$, called the adjoint, satisfying

$$
\left\langle\mathcal{A}^{*} y, x\right\rangle_{\mathbf{X}}=\langle y, \mathcal{A} x\rangle_{\mathbf{Y}} \quad \text { for all points } \quad x \in \mathbf{X}, y \in \mathbf{Y}
$$

When $\mathbf{X}=\mathbf{R}^{n}$ and $\mathbf{Y}=\mathbf{R}^{m}$, every linear map $\mathcal{A}$ can be identified with a matrix $A \in \mathbf{R}^{m \times n}$. In this case, the matrix associated with the adjoint $\mathcal{A}^{*}$ is the transpose $A^{T}$.

Note: The adjoint differs significantly from the classical adjoint in Cramer's Rule.

## Self-adjoint Linear Operators

Let $(\mathbf{E},\langle\cdot, \cdot\rangle)$ be a Euclidean space and let $\mathcal{A} \in \mathbf{L}(\mathbf{E}, \mathbf{E})$.
We say that $\mathcal{A}$ is self-adjoint if $\mathcal{A}=\mathcal{A}^{*}$. The set of all self-adjoint linear operators on $\mathbf{E}$ is denoted by $\mathcal{S}(\mathbf{E})$ or $\mathcal{S}(\mathbf{E},\langle\cdot, \cdot\rangle)$ if great specificity is required.

If $\mathbf{E}=\mathbf{R}^{n}$, the matrix representation of a self-adjoint linear operator is a symmetric matrix.

A self-adjoint linear operator on $\mathbf{R}^{n}$ can be identified with the symmetric matrices on $\mathbf{R}^{n}$ and so form a subspace of $\mathbf{R}^{n \times n}$ which we denote by $\mathbf{S}^{n}:=\left\{A \in \mathbf{R}^{n \times n} \mid A=A^{T}\right\}$.

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Example: Let $A \in \mathbf{S}^{n}$ and define $\mathcal{H}: \mathbf{S}^{n} \rightarrow \mathbf{S}^{n}$ by $\mathcal{H}(X):=A X A$. Then $\mathcal{H} \in \mathcal{S}\left(\mathbf{S}^{n}\right)$ is a self-adjoint linear operator.

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## Positive Semi-Definite Linear Operators

A self-adjoint operator $\mathcal{A}$ is positive semi-definite, denoted $\mathcal{A} \succeq 0$, whenever

$$
\langle\mathcal{A} x, x\rangle \geq 0 \quad \text { for all } x \in \mathbf{E}
$$

Similarly, a self-adjoint operator $\mathcal{A}$ is positive definite, denoted $\mathcal{A} \succ 0$, whenever

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\langle\mathcal{A} x, x\rangle>0 \quad \text { for all } 0 \neq x \in \mathbf{E}
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- A bilinear form $b(\cdot, \cdot)$ on the Euclidean space $(\mathbf{E},\langle\cdot, \cdot\rangle)$ is an inner product on $\mathbf{E}$ if and only if there is a positive definite linear operator $\mathcal{A}$ on $\mathbf{E}$ such that $b(x, y)=\langle\mathcal{A} x, y\rangle \forall x, y \in \mathbf{E}$.


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- For any two linear operators $\mathcal{A}$ and $\mathcal{B}$, we will use the notation $\mathcal{A} \succeq \mathcal{B}$ to mean $\mathcal{A}-\mathcal{B} \succeq 0$. The notation $\mathcal{A} \succ \mathcal{B}$ is defined similarly.


## Norms

A norm on a vector space $\mathcal{V}$ is a function $\|\cdot\|: \mathcal{V} \rightarrow \mathbf{R}$ for which the following three properties hold for all point $x, y \in \mathcal{V}$ and scalars $a \in \mathbf{R}$ :
(Absolute homogeneity) $\|a x\|=|a| \cdot\|x\|$
(Triangle inequality) $\|x+y\| \leq\|x\|+\|y\|$
(Positivity) Equality $\|x\|=0$ holds if and only if $x=0$.

The inner product in the Euclidean space $\mathbf{E}$ always induces a norm $\|x\|=\sqrt{\langle x, x\rangle}$. Unless specified otherwise, the symbol $\|x\|$ for $x \in \mathbf{E}$ will always denote this induced norm.

## Examples of Norms

p-norms on $\mathbf{R}^{n}$ :

$$
\|x\|_{p}= \begin{cases}\left(\left|x_{1}\right|^{p}+\ldots+\left|x_{n}\right|^{p}\right)^{1 / p} & \text { for } 1 \leq p<\infty \\ \max \left\{\left|x_{1}\right|, \ldots,\left|x_{n}\right|\right\} & \text { for } p=\infty\end{cases}
$$

Elliptic or inner product norms on $\mathbf{R}^{n}$ : Let $\mathcal{A} \in \mathbf{L}(\mathbf{E}, \mathbf{E})$ be positive definite.

$$
\|x\|_{\mathcal{A}}:=\sqrt{\langle A x, y\rangle}
$$

Dual norms: Given an arbitrary norm $\|\cdot\|$ on $\mathbf{R}^{n}$, the norm dual to $\|\cdot\|$ is defined by

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Why do norms and their duals satisfy the generalized Cauchy-Schwartz inequality

$$
|\langle x, y\rangle| \leq\|x\| \cdot\|y\|^{*} \quad \text { for all } x, y \in \mathbf{E} ?
$$

## Equivalence of Norms

All norms on $\mathbf{E}$ are "equivalent" in the sense that for any two norms $\rho_{1}(\cdot)$ and $\rho_{2}(\cdot)$, there exist constants $\alpha, \beta>0$ satisfying

$$
\alpha \rho_{1}(x) \leq \rho_{2}(x) \leq \beta \rho_{1}(x) \quad \text { for all } x \in \mathbf{E} .
$$

$$
\begin{aligned}
\|x\|_{2} & \leq\|x\|_{1} \leq \sqrt{n}\|x\|_{2} \\
\|x\|_{\infty} & \leq\|x\|_{2} \leq \sqrt{n}\|x\|_{\infty} \\
\|x\|_{\infty} & \leq\|x\|_{1} \leq n\|x\|_{\infty} .
\end{aligned}
$$

The term "equivalent" is a misnomer since the constants $\alpha, \beta$ strongly depend on the (often enormous) dimension of the vector space $\mathbf{E}$. Hence measuring quantities in different norms can yield strikingly different conclusions.

## The Orthogonal Group

Let $(\mathbf{E},\langle\cdot, \cdot\rangle)$ be a Euclidean space. A linear operator $U \in \mathbf{L}(\mathbf{E}, \mathbf{E})$ is said to be distance preserving if

$$
\|U x\|=\|x\| \forall x \in \mathbf{E}
$$

where $\|x\|=\sqrt{\langle x \cdot x\rangle}$ is the inner product norm on $\mathbf{E}$. The set $\mathcal{O}(\mathbf{E})$ of all distance preserving linear operators on $\mathbf{E}$ is called the orthogonal group for $\mathbf{E}$, and the elements of $\mathcal{O}(\mathbf{E})$ are called orthogonal operators.

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- Given a basis for $\mathbf{E}$, we can identify $\mathbf{L}(\mathbf{E}, \mathbf{E})$ with $\mathbf{R}^{n \times n}$ where $n$ is the dimension of $\mathbf{E}$. If we identify $\mathcal{O}(\mathbf{E})$ with its associated matrices, then $\mathcal{O}(\mathbf{E})=\left\{U \in \mathbf{R}^{n \times n} \mid U U^{T}=I=U^{T} U\right\}$ and its elements are called orthogonal matrices.


## Eigenvalues of Symmetric Matrices

Let $A \in \mathbf{S}^{n} . \lambda \in \mathbf{R}$ is an eigenvalue for $A$ if exists $x \in \mathbf{R}^{n} \backslash\{0\}$ s.t. $A x=\lambda x$.

The vector $x$ is called an eigenvector associated with $\lambda$.

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If $A \in \mathbf{S}^{n}$, these $n$ roots are real. One can show that there is an associated orthonormal basis of real eigenvectors. Consequently, $A$ is diagonalizable in the sense that

$$
U^{T} A U=\Lambda \quad \text { or } \quad A=U \Lambda U^{T}
$$

where the columns of $U \in \mathcal{O}\left(\mathbf{R}^{n}\right)$ are an orthonormal basis of eigenvectors and $\Lambda$ is the diagonal matrix of corresponding eigenvalues.

## Rayleigh-Ritz Theorem and Square Roots

Fix an ordering and denote the eigenvalues of $A$ by

$$
\lambda_{1}(A) \geq \lambda_{2}(A) \geq \ldots \geq \lambda_{n}(A)
$$

A simple consequence of the decomposition $A=U \Lambda U^{T}$ is the Rayleigh-Ritz theorem:

$$
\lambda_{n}(A) \leq \frac{\langle A u, u\rangle}{\langle u, u\rangle} \leq \lambda_{1}(A) \quad \text { for all } u \in \mathbf{R}^{n} \backslash\{0\}
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Observe that the two conditions, $A \succeq 0$ and $\lambda_{n}(A) \geq 0$ are equivalent; similarly, $A \succ 0$ if and only $\lambda_{n}(A)>0$.

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Consequently, $A \in \mathbf{S}^{n}$ is positive semidefinite if and only if there exists a matrix $B \in \mathbf{S}^{n}$ satisfying $A=B B^{T}$ (why?). The matrix $B$ is called a square root of $A$. There are infinitely many such such square roots (see Cholesky Factorizations). The spectral square root is $B=U \Lambda^{1 / 2} U^{T}=: \sqrt{A}$.

## The Singular Value Decomposition

Given $A, B^{T} \in \mathbf{R}^{m \times n}$, one can show that the nonzero eigenvalues of $A B$ coincide with those of $B A$ including multiplicity.

Therefore, the eigenvalues of the symmetric matrices $A^{T} A$ and $A A^{T}$ coincide up to multiplicity. Since these matrices are positive semi-definite (why?), their nonzero eigenvalues are positives and coincide up to multiplicity.

Let $k:=\min \{n, m\}$ and define

$$
\sigma_{1}(A) \geq \sigma_{2}(A) \geq \ldots \geq \sigma_{k}(A) \geq 0
$$

to be the largest $k$ eigenvalues of $\sqrt{A^{T} A}$ and note that any other eigenvalue of $\sqrt{A^{T} A}$ must be zero. The $\sigma_{i}$ s are called the singular values of $A$.

## The Singular Value Decomposition

If the columns of $V \in \mathcal{O}\left(\mathbf{R}^{n}\right)$ form an orthonormal basis of eigenvectors for $A^{T} A$ ordered in correspondence with the magnitude of its eigenvalues, it can be shown that there is a corresponding $U \in \mathcal{O}\left(\mathbf{R}^{m}\right)$ whose columns form an orthonormal basis of eigenvectors for $A A^{T}$ such that

$$
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where the principal diagonal $\Sigma \in \mathbf{R}^{m \times n}$ are the ordered singular values of $A$ with all other values zero.

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If we let $k:=\operatorname{rank}(A)$, we may write

$$
A=U \Sigma V^{T}
$$

where now $U \in \mathbf{R}^{m \times k}, V \in \mathbf{R}^{n \times k}$ have orthogonal columns and $\Sigma \in \mathbf{R}^{k \times k}$ is diagonal with the ordered nonzero singular values on the diagonal. This called the compact or reduced singular value decomposition.

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For this reason, some authors refer to only the nonzero singular values as the singular values. The columns of $U$ are called the left singular vectors and those of $V$ are the right singular vectors.

## The Operator Norm on $\mathbf{R}^{n \times n}$

The Rayleigh-Ritz Theorem tells us that

$$
\|A\|_{\mathrm{op}}:=\sup _{x:\|x\| \leq 1}\|A x\|=\sigma_{1}(A)
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Let $\sigma: \mathbf{R}^{m \times n} \rightarrow \mathbf{R}^{k}$, where $k:=\min \{m, n\}$, be the mapping that takes a matrix to its ordered vector of singular values:

$$
\sigma(A):=\left(\sigma_{1}(A), \sigma_{2}(A), \ldots, \sigma_{k}(A)\right)^{T}
$$

The Schatten p-norm of a $A \in \mathbf{R}^{m \times n}$, for $1 \leq p \leq \infty$ is given by

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\|A\|_{p}:=\|\sigma(A)\|_{p}
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Hence $\|A\|_{\mathrm{op}}=\|\sigma(A)\|_{\infty}$. For $p=1,\|A\|_{1}$ is called the nuclear or trace norm.

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It can be shown that all of the Schatten $p$-norms are norms on the Euclidean space $\mathbf{R}^{m \times n}$.

## Sets and Operations on Sets

Let $\mathbf{X}, \mathbf{Y}$ be Euclidean spaces with $X_{i} \subset \mathbf{X} i=1,2, Y \subset \mathbf{Y}$, and let $\mathcal{A} \in \mathbf{L}(\mathbf{X}, \mathbf{Y})$.
$-\mathbf{R}_{+}^{n}:=\left\{x \in \mathbf{R}^{n} \mid x_{i} \geq 0, i=1, \ldots, n\right\}, \mathbf{R}_{++}^{n}:=\left\{x \in \mathbf{R}^{n} \mid x_{i}>0, i=1, \ldots, n\right\}$
$-\mathbf{S}^{n}{ }_{+}:=\left\{H \in \mathbf{S}^{n} \mid H \succeq 0\right\}, \quad \mathbf{S}^{n}{ }_{++}:=\left\{H \in \mathbf{S}^{n} \mid H \succ 0\right\}$

- For $\lambda \in \mathbf{R}, \lambda X:=\{\lambda x \mid x \in X\}$.
$-X_{1}+X_{2}:=\left\{x^{1}+x^{2} \mid x_{i} \in X_{i}, i=1,2\right\}$ with $X_{1}-X_{2}$ defined similarly.
$-\mathbf{R}_{+} Y:=\left\{\lambda y \mid \lambda \in \mathbf{R}_{+}, y \in Y\right\}$, the cone generated by $Y$.
- An affine set is a translate of a subspace.
- The affine hull of $Y$, aff $Y$, is the intersection of all affine sets containing $Y$.
$-\mathcal{A} X_{1}:=\left\{\mathcal{A} x \mid x \in X_{1}\right\}$
$-\mathcal{A}^{-1} Y:=\{x \mid \mathcal{A} x \in Y\}$
- $X_{1} \times Y:=\left\{(x, y) \mid x \in X_{1}, y \in Y\right\}$


## Convex Sets

A set $C \subset \mathbf{E}$ is said to be convex if

$$
x, y \in C \text { and } \lambda \in[0,1] \quad \Longrightarrow \quad(1-\lambda) x+\lambda y \in C .
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That is, $C$ contains all line segments connecting points in $C$.
Let $\mathcal{A} \in \mathbf{L}(\mathbf{E}, \mathbf{Y})$. If $C_{1}, C_{2} \subset \mathbf{E}$ and $K \subset \mathbf{Y}$ are all convex, then so are the sets
$-C_{1} \cap C_{2}$
$-\mathbf{R}_{+} K$ and $\lambda K \quad \forall \lambda \in \mathbf{R}$
$-C_{1}+C_{2}$
$-\mathcal{A} C_{1}$ and $\mathcal{A}^{-1} K$

- $C_{1} \times K$
- cl $K$ and intr $K$


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That is, $C$ contains all line segments connecting points in $C$.
Let $\mathcal{A} \in \mathbf{L}(\mathbf{E}, \mathbf{Y})$. If $C_{1}, C_{2} \subset \mathbf{E}$ and $K \subset \mathbf{Y}$ are all convex, then so are the sets
$-C_{1} \cap C_{2}$
$-\mathbf{R}_{+} K$ and $\lambda K \quad \forall \lambda \in \mathbf{R}$
$-C_{1}+C_{2}$
$-\mathcal{A} C_{1}$ and $\mathcal{A}^{-1} K$

- $C_{1} \times K$
- cl $K$ and intr $K$

We will spend a lot of time with convex sets.

## Point-Set Topology

Let $\mathbf{E}$ be a Euclidean space with $x \in X \subset \mathbf{E}$.

- Given $r>0$, the open $r$ ball around $x$ is the set

$$
B_{r}(x):=\{y \mid\|x-y\|<r\} .
$$

$-x$ is in the closure of $X$, written $x \in \operatorname{cl} X$, if

$$
B_{r}(x) \cap X \neq \emptyset \forall r>0 .
$$

- $X$ is closed if $X=\operatorname{cl} X$.
$-x \in X$ is in the interior of $X$, written $x \in \operatorname{intr} X$, if there is an $r>0$ such that $B_{r}(x) \subset X$.
$-X$ is open if $X=\operatorname{intr} X$.
$-X$ is bounded if there is an $r>0$ such that $X \subset B_{r}(0)$.
$-X$ is compact if it is closed and bounded.


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- $X$ is compact if it is closed and bounded.

Theorem (Bolzano-Weierstrass)
$Q \subset \mathbf{E}$ is compact if and only if every sequence in $Q$ admits a subsequence converging to a point in $Q$.

## Limits Inferior and Superior

Define the extended real line $\overline{\mathbf{R}}:=\mathbf{R} \cup\{ \pm \infty\}$.
The limit inferior and limit superior of any sequence $\left\{r_{i}\right\} \subset \overline{\mathbf{R}}$ are defined by
$\liminf _{i \rightarrow \infty} r_{i}=\lim _{i \rightarrow \infty}\left\{\inf _{j \geq i} r_{j}\right\} \quad$ and $\quad \limsup _{i \rightarrow \infty} r_{i}=\lim _{i \rightarrow \infty}\left\{\sup _{j \geq i} r_{j}\right\}$.

For any function $f: \mathbf{E} \rightarrow \overline{\mathbf{R}}$ and a point $x \in \mathbf{E}$, we set

$$
\liminf _{y \rightarrow x} f(y)=\lim _{r>0}\left\{\inf _{y \in B_{r}(x) \backslash\{x\}} f(y)\right\}
$$

The symbol $\limsup _{y \rightarrow x} f(y)$ is defined similarly, with sup replacing inf.

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Note: The infimum (supremum) over the empty set is $+\infty(-\infty)$.

## Functions and Continuity

Let $f: \mathbf{E} \rightarrow \overline{\mathbf{R}}$ and $F: \mathbf{X} \rightarrow \mathbf{Y}$.
$-\operatorname{dom} f:=\{x \mid f(x)<\infty\}$

- epi $f:=\{(x, \lambda) \mid f(x) \leq \lambda\} \subset \mathbf{E} \times \mathbf{R}$
$-f$ is lower semi-continuous (lsc) at $x \in \mathbf{E}$ if $\liminf _{y \rightarrow x} f(y) \geq f(x) . f$ is closed if it is lsc for all $x \in \mathbf{E}$.
$-f$ is upper semi-continuous at $x \in \mathbf{E}$ if $f(x) \geq \limsup _{y \rightarrow x} f(y)$.
$-f$ is continuous at $x \in \operatorname{intr}(\operatorname{dom} f)$ if $\liminf _{y \rightarrow x} f(y)=f(x)=\limsup _{y \rightarrow x} f(y)$.
$-F$ continuous at $x \in \mathbf{X}$ if

$$
\forall \epsilon>0 \exists \delta>0 \text { s.t. }\|F(y)-F(x)\| \leq \epsilon \text { when }\|y-x\| \leq \delta .
$$

- For $L>0, F$ is L-Lipschitz continuous at $x \in \mathbf{X}$ if

$$
\|F(x)-F(y)\| \leq L\|x-y\| .
$$

- For $L>0$ and $X \subset \mathbf{X}, F$ is L-Lipschitz continuous on $X$ if it is L-Lipschitz continuous forall $x \in \mathbf{X}$. If $X=\mathbf{X}$, we simply say $F$ is L-Lipschitz. If $0<L<1$, we say that $F$ is a contraction.


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Theorem
$f: \mathbf{E} \rightarrow \overline{\mathbf{R}}$ is closed if and only if epi $f$ is closed.

## Existence of Optimal Solutions

Theorem (Weierstrass Extrema Value Theorem)
A continuous function on a compact set attains its extrema values on that set. That is, if $f: C \rightarrow \mathbf{R}$ is continuous on the compact set $C \subset \mathbf{E}$, then there exist $\bar{x}, \bar{y} \in C$ such that $f(\bar{x}) \leq f(x) \leq f(\bar{y})$ for all $x \in C$.
This can be refined using lower semi-continuity.
Theorem
If $f: Q \rightarrow \mathbf{R}$ is closed with $Q \subset \mathbf{E}$ compact, then there is an $\bar{x} \in Q$ such that $f(\bar{x}) \leq f(x)$ for all $x \in Q$.

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Coercive Functions: A function $f: \mathbf{E} \rightarrow \overline{\mathbf{R}}$ is coercive if for any sequence $x_{i}$ with $\left\|x_{i}\right\| \rightarrow \infty$, it must be that $f\left(x_{i}\right) \rightarrow+\infty$.
It is easy to show that $f$ is coercive if and only if the sets $\{x \mid f(x) \leq r\}$ are compact for all $r \in \mathbf{R}$. This observation implies that any closed coercive function has a global minimizer, i.e. there is $\bar{x}$ such that $f(\bar{x}) \leq f(x)$ for all $x \in \mathbf{E}$.

## Linear Operators

Let $\mathbf{X}$ and $\mathbf{Y}$ be real normed linear spaces with norms $\|\cdot\|_{x}$ and $\|\cdot\|_{y}$, respectively.
A linear transformation (or operator) from $\mathbf{X}$ to $\mathbf{Y}$ is any mapping $\mathcal{L}: \mathbf{X} \rightarrow \mathbf{Y}$ such that

$$
\mathcal{L}(\alpha x+\beta z)=\alpha \mathcal{L}(x)+\beta \mathcal{L}(z) \quad \forall x, z \in \mathbf{X} \text { and } \alpha, \beta \in \mathbf{R} .
$$

The linear operator $\mathcal{T}$ is continuous with respect to the norms on $\mathbf{X}$ and $\mathbf{Y}$ if and only if

$$
\|\mathcal{T}\|:=\sup _{\|x\|_{x} \leq 1}\|\mathcal{T} x\|_{y} \quad \forall \mathcal{T} \in \mathbf{L}[\mathbf{X}, \mathbf{Y}]
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is finite.
Let $\mathbf{L}[\mathbf{X}, \mathbf{Y}]$ denote the space of all continuous linear operators from $\mathbf{X}$ to $\mathbf{Y}$. In can be shown that $\|\mathcal{T}\|$ is a norm on this space.

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is finite.
Let $\mathbf{L}[\mathbf{X}, \mathbf{Y}]$ denote the space of all continuous linear operators from $\mathbf{X}$ to $\mathbf{Y}$. In can be shown that $\|\mathcal{T}\|$ is a norm on this space.
The topological dual of the normed linear space $\mathbf{X}$ is

$$
\mathbf{X}^{*}:=\mathbf{L}[\mathbf{X}, \mathbf{R}]
$$

with the duality pairing denoted by

$$
\langle\phi, x\rangle=\phi(x) \quad \forall(\phi, x) \in \mathbf{X}^{*} \times \mathbf{X}
$$

## Hilbert Spaces

If the norm on $\mathbf{X}$ satisfies the parallelogram law,

$$
\|x-y\|^{2}+\|x+y\|^{2}=2\|x\|^{2}+2\|y\|^{2}
$$

then we call $\mathbf{X}$ a Hilbert space.

In this case there of a natural isometry between $\mathbf{X}^{*}$ and $\mathbf{X}$ under which the duality pairing is an inner product:

$$
\langle x, y\rangle=\frac{\|x+y\|^{2}-\|x-y\|^{2}}{4} .
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$$

Note: A Euclidean space is a real finite dimensional Hilbert space.

## Bilinear Forms

Let $\mathbf{X}$ and $\mathbf{Y}$ be real linear spaces. A mapping $\mathcal{Q}: \mathbf{X} \times \mathbf{X} \rightarrow \mathbf{Y}$ is said to be a bilinear if it is linear in each argument separately: for all $\left(x^{i}, z^{j}\right) \in \mathbf{X} \times \mathbf{X}, i=1,2$, and $\alpha, \beta, \gamma, \delta \in \mathbf{R}$

$$
\begin{aligned}
\mathcal{Q}\left(\alpha x^{1}+\beta x^{2}, \gamma z^{1}+\delta z^{2}\right) & =\alpha \mathcal{Q}\left(x^{1}, \gamma z^{1}+\delta z^{2}\right)+\beta \mathcal{Q}\left(x^{2}, \gamma z^{1}+\delta z^{2}\right) \\
& =\gamma \mathcal{Q}\left(\alpha x^{1}+\beta x^{2}, z^{1}\right)+\delta \mathcal{Q}\left(\alpha x^{1}+\beta x^{2}, z^{2}\right)
\end{aligned}
$$

The bilinear form $\mathcal{Q}$ is said to be symmetric if $\mathcal{Q}(x, z)=\mathcal{Q}(z, x)$.

Let $\mathbf{B}[\mathbf{X}, \mathbf{Y}]$ denote the set of all continuous bilinear maps from $\mathbf{X}$ to $\mathbf{Y}$.

If $\mathbf{Y}=\mathbf{R}$, the bilinear map $\mathcal{Q}$ is call a bilinear form and we write $\mathbf{Q}[\mathbf{X}]:=\mathbf{B}[\mathbf{X}, \mathbf{R}]$.

## Differentiability

Let $U \subset \mathbf{E}$ be open.
$f: U \rightarrow \mathbf{R}$ is differentiable at $x \in U$ if there exists a vector, denoted by $\nabla f(x) \in \mathbf{E}$, satisfying

$$
\lim _{h \rightarrow 0} \frac{f(x+h)-f(x)-\langle\nabla f(x), h\rangle}{\|h\|}=0 .
$$

We call $\nabla f(x)$ the gradient of $f$ at $x$. If $\mathbf{E}=\mathbf{R}^{n}$,

$$
\nabla f(x)=\left(\begin{array}{c}
\frac{\partial f(x)}{\partial x_{1}} \\
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\frac{\partial f(x)}{\partial x_{n}}
\end{array}\right)
$$

Let the symbol $o(r)$ represent the class of functions satisfying $0=\lim _{r \downarrow 0} o(r) / r$. Then $f$ is differentiable at $x$ if and only if

$$
f(x+h)=f(x)+\langle\nabla f(x), h\rangle+o(\|h\|)
$$

## Differentiability

If the mapping $\nabla f: U \rightarrow \mathbf{R}^{n}$ is well-defined and continuous, we say $f$ is $\mathcal{C}^{1}$-smooth on $U$.

If the gradient satisfies the stronger Lipschitz property

$$
\|\nabla f(y)-\nabla f(x)\| \leq \beta\|y-x\| \quad \text { holds for all } x, y \in U
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More generally, a mapping $F: U \rightarrow \mathbf{Y}$ is differentiable at $x \in U$ if there exists a linear mapping from $\mathbf{E}$ to $\mathbf{Y}$, denoted by $F^{\prime}(x)$, satisfying

$$
F(x+h)=F(x)+F^{\prime}(x) h+o(\|h\|) .
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$$
F(x+h)=F(x)+F^{\prime}(x) h+o(\|h\|) .
$$

If one chooses bases in $\mathbf{E}$ and $\mathbf{Y}$, then $F^{\prime}(x) \in \mathbf{L}(\mathbf{E}, \mathbf{Y})$ can be given a matrix representation which is denoted by $\nabla F(x)$ and called the Jacobian of $F$ at $x$. If the assignment $x \mapsto F^{\prime}(x)$ is continuous, we say that $F$ is $\mathcal{C}^{1}$-smooth.

## Differentiability

If $\mathbf{E}=\mathbf{R}^{n}$ and $\mathbf{Y}=\mathbf{R}^{m}$, we can write $F$ in terms of coordinate functions $F(x)=\left(F_{1}(x), \ldots, F_{m}(x)\right)$, and then the Jacobian is simply

$$
\nabla F(x)=\left(\begin{array}{c}
\nabla F_{1}(x)^{T} \\
\nabla F_{2}(x)^{T} \\
\vdots \\
\nabla F_{m}(x)^{T}
\end{array}\right)=\left(\begin{array}{cccc}
\frac{\partial F_{1}(x)}{\partial x_{1}} & \frac{\partial F_{1}(x)}{\partial x_{2}} & \ldots & \frac{\partial F_{1}(x)}{\partial x_{n}} \\
\frac{\partial F_{2}(x)}{\partial x_{1}} & \frac{\partial F_{2}(x)}{\partial x_{2}} & \ldots & \frac{\partial F_{2}(x)}{\partial x_{n}} \\
\vdots & \vdots & \ddots & \vdots \\
\frac{\partial F_{m}(x)}{\partial x_{1}} & \frac{\partial F_{m}(x)}{\partial x_{2}} & \ldots & \frac{\partial F_{m}(x)}{\partial x_{n}}
\end{array}\right) .
$$

## Calculus Rules

Let $U \subset \mathbf{E}$ and $W \subset \mathbf{Y}$ be open.
Let $F_{i}: U \rightarrow \mathbf{Y}, i=1,2, F: U \rightarrow W$, and $H: W \rightarrow \mathbf{Z}$ be $\mathcal{C}^{1}$ (this can be significantly weakened).

- If $F \in \mathbf{L}(\mathbf{E}, \mathbf{Y})$, the $F^{\prime}(x)=F$ for al $x \in \mathbf{E}$.
- For all $\lambda \in \mathbf{R}$ and $x \in U,^{\prime}(\lambda F)^{\prime}(x)=\lambda F^{\prime}(x)$.
- For all $x \in U,\left(F_{1}+F_{2}\right)^{\prime}(x)=F_{1}^{\prime}(x)+F_{2}^{\prime}(x)$.
- The Chain Rule: The mapping $G: U \rightarrow \mathbf{Z}$ given by
$G:=H \circ F$ is differentiable on $U$ with $G^{\prime}(x)=H^{\prime}(F(x)) \circ F^{\prime}(x)$.


## Example

Let $A \in \mathbf{R}^{s \times n}$ and $B \in \mathbf{R}^{n \times t}$ and consider the mapping $\mathcal{T}: \mathbf{R}^{m \times n} \rightarrow \mathbf{R}^{s \times t}$ given by

$$
\mathcal{T}(X):=A X B .
$$

Clearly, $\mathcal{T} \in \mathbf{L}\left(\mathbf{R}^{m \times n}, \mathbf{R}^{s \times t}\right)$, hence

$$
\mathcal{T}^{\prime}(X) Y=\mathcal{T}(Y)=A Y B \quad \forall X \in \mathbf{R}^{m \times n}
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What is $\nabla \mathcal{T}$ ?

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What is $\nabla \mathcal{T}$ ?

Representing the matrix $\nabla \mathcal{T}$ requires choosing bases in both $\mathbf{R}^{m \times n}$ and $\mathbf{R}^{s \times t}$ and then recording the action of $\mathcal{T}$ on these bases. This is doable, but it is a real mess. A helpful tool in this regard is the Kronecker product to be discussed later.

## The Second Derivative

Let $F: \mathbf{X} \rightarrow \mathbf{Y}$ we say that $F$ is twice differentiable at $x$ if $F$ is differentiable at $x$ and there is a bilinear form $\mathcal{Q}$ such that

$$
\lim _{z \rightarrow x} \frac{\left\|F(z)-\left(F(z)+\nabla F(x)(z-x)+\frac{1}{2} \mathcal{Q}(z-x, z-x)\right)\right\|}{\|y-x\|^{2}}=0
$$

We call $\mathcal{Q}$ the second derivative of $F$ at $x$ and write $\mathcal{Q}=F^{\prime \prime}(x)$. If the mapping $x \rightarrow F^{\prime \prime}(x)$ is continuous, we say that $F$ is $\mathcal{C}^{2}$.

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When $\mathbf{X}=\mathbf{R}^{n}$ and $\mathbf{Y}=\mathbf{R}$, we call $F^{\prime \prime}(x)$ the Hessian of $F$ at $x$ and write $\nabla^{2} F(x):=F^{\prime \prime}(x)$. If all of the second partials of $F$ are continuous, then $\nabla^{2} F(x) \in \mathbf{S}^{n}$ is the $n \times n$ matrix of second partials.

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Again, the little-o notation gives

$$
F(y)=F(x)+\langle\nabla F(x),(y-x)\rangle+\frac{1}{2}\left\langle\nabla^{2} F(x)(y-x),(y-x)\right\rangle+o\left(\|y-x\|^{2}\right)
$$

## Computing Derivatives

Consider the linear transformation $\mathcal{T} \in \mathbf{L}\left[\mathbf{R}^{n \times n}, \mathbf{R}^{n \times n}\right]$ given by

$$
\mathcal{T}(X)=A X+X B \quad \text { for fixed } A, B \in \mathbf{R}^{n \times n}
$$

and let $F: \mathbf{R}^{n} \rightarrow \mathbf{R}^{n \times n}$ be given by

$$
F(x):=\operatorname{diag}(x),
$$

where the linear transformation $\operatorname{diag}(\cdot) \in \mathbf{L}\left[\mathbf{R}^{n}, \mathbf{R}^{n \times n}\right]$ maps $x$ to the $n \times n$ matrix whose diagonal is $x$. What is $(\mathcal{T} \circ \operatorname{diag})^{\prime}(x)$ ?

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Since both $\mathcal{T}$ and diag are linear, so is $(\mathcal{T} \circ \operatorname{diag})$. Therefore,

$$
(\mathcal{T} \circ \operatorname{diag}(\cdot))^{\prime}(x)(d)=(\mathcal{T} \circ \operatorname{diag}(\cdot))(d)=A \operatorname{diag}(d)+\operatorname{diag}(d) B
$$ for all $x \in \mathbf{R}^{n}$.

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F(x):=\operatorname{diag}(x),
$$

where the linear transformation $\operatorname{diag}(\cdot) \in \mathbf{L}\left[\mathbf{R}^{n}, \mathbf{R}^{n \times n}\right]$ maps $x$ to the $n \times n$ matrix whose diagonal is $x$. What is $(\mathcal{T} \circ \operatorname{diag})^{\prime}(x)$ ?

Since both $\mathcal{T}$ and diag are linear, so is $(\mathcal{T} \circ \operatorname{diag})$. Therefore,

$$
(\mathcal{T} \circ \operatorname{diag}(\cdot))^{\prime}(x)(d)=(\mathcal{T} \circ \operatorname{diag}(\cdot))(d)=A \operatorname{diag}(d)+\operatorname{diag}(d) B
$$ for all $x \in \mathbf{R}^{n}$.

What is $\nabla(\mathcal{T} \circ \operatorname{diag}(\cdot))$ ?

## Computing Derivatives

Let $A \in \mathbf{R}^{m \times n}, b \in \mathbf{R}^{m}$, and define $f: \mathbf{R}^{n} \rightarrow \mathbf{R}$ by

$$
f(x):=\frac{1}{2}\|A x-b\|^{2} .
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Compute $\nabla f(x)$ and $\nabla^{2} f(x)$.

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$$
\begin{aligned}
f(x+\Delta x) & =\frac{1}{2}\|(A x-b)+A \Delta x\|^{2} \\
& =\frac{1}{2}\|A x-b\|^{2}+\langle A x-b, A \Delta x\rangle+\frac{1}{2}(\Delta x)^{T} A^{T} A \Delta x \\
& =f(x)+\left\langle A^{T}(A x-b), \Delta x\right\rangle+\frac{1}{2}\left\langle\left(A^{T} A\right) \Delta x, \Delta x\right\rangle .
\end{aligned}
$$

Therefore,

$$
\nabla f(x)=A^{T}(A x-b) \quad \text { and } \quad \nabla^{2} f(x)=A^{T} A
$$

## Computing Derivatives

Let $A \in \mathbf{R}^{m \times n}, B \in \mathbf{R}^{n \times n}$, and $C \in \mathbf{R}^{n \times k}$, and define $\mathcal{Q}: \mathbf{R}^{n \times n} \times \mathbf{R}^{n \times n} \rightarrow \mathbf{R}^{m \times k}$ by

$$
\mathcal{Q}(X, Z)=A X^{T} B Z C .
$$

$\mathcal{Q}$ is a bilinear mapping in $\mathbf{B}\left[\mathbf{R}^{n \times n}, \mathbf{R}^{m \times k}\right]$. This bilinear mapping is a bilinear form if $m=k=1$, and it is symmetric if $m=k=1$, $A^{T}=C$, and $B \in \mathbf{S}^{n}$.
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$$
\begin{aligned}
\mathcal{Q}(X & +\Delta X, Z+\Delta Z) \\
& =A(X+\Delta X)^{T} B(Z+\Delta Z) C \\
& =A X^{T} B Z C+A(\Delta X)^{T} B Z C+A X B(\Delta Z) C+A(\Delta X)^{T} B(\Delta Z) C \\
& =\mathcal{Q}(X, Z)+\left(A(\Delta X)^{T} B Z C+A X B(\Delta Z) C\right)+\frac{1}{2}(2 \mathcal{Q}(\Delta X, \Delta Z)) .
\end{aligned}
$$

$\mathcal{Q}^{\prime}(X, Z)(U, V)=\mathcal{Q}(U, Z)+\mathcal{Q}(X, V)$ and $\mathcal{Q}^{\prime \prime}(X, Z)(U, V)=2 \mathcal{Q}(U, V)$.

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Is this true of all bilinear forms regardless of the space?

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Is this true of all bilinear forms regardless of the space?
What is the gradient and Hessian when $m=k=1, A^{T}=C$, and $B \in \mathbf{S}^{n}$ ?

## Accuracy of Linear and Quadratic Approximations

Let $U \subset \mathbf{E}$ be open. Consider a function $f: U \rightarrow \mathbf{R}$ and a point $x \in U$. Multivariate calculus identifies the following two functions as the "best" linear and quadratic approximations of $f$ near $x$, respectively:

$$
\begin{aligned}
l_{x}(y) & :=f(x)+\langle\nabla f(x), y-x\rangle \\
Q_{x}(y) & :=f(x)+\langle\nabla f(x), y-x\rangle+\frac{1}{2}\left\langle\nabla^{2} f(x)(y-x), y-x\right\rangle
\end{aligned}
$$

Can we quantify how well these functions approximate $f$ near $x$ ?

## Accuracy of Linear and Quadratic Approximations

Given $x, y \in \mathbf{E}$ define $\varphi: \mathbf{R} \rightarrow \mathbf{R}$ by

$$
\varphi(t):=f(x+t(y-x)) .
$$

Then the following approximation results follow directly from Taylor approximations to $\varphi$ since $\varphi^{\prime}(0)=\langle\nabla f(x), y-x\rangle$ and $\varphi^{\prime \prime}(0)=\left\langle\nabla^{2} f(x)(y-x), y-x\right\rangle$.
Theorem (Accuracy in approximation)
Consider a $C^{1}$-smooth function $f: U \rightarrow \mathbf{R}$ and two points $x, y \in U$. Then we have

$$
f(y)=l_{x}(y)+\int_{0}^{1}\langle\nabla f(x+t(y-x))-\nabla f(x), y-x\rangle d t
$$

If $f$ is $C^{2}$-smooth, then the equation holds:

$$
f(y)=Q_{x}(y)+\int_{0}^{1} \int_{0}^{t}\left\langle\left(\nabla^{2} f(x+s(y-x))-\nabla^{2} f(x)\right)(y-x), y-x\right\rangle d s d t
$$

## Accuracy of Linear and Quadratic Approximations

Corollary (Accuracy in approximation under Lipschitz conditions)

1 Suppose that $f: U \rightarrow \mathbf{R}$ is a $\beta$-smooth function. Then for any points $x, y \in U$ the inequality

$$
\left|f(y)-l_{x}(y)\right| \leq \frac{\beta}{2}\|y-x\|^{2} \quad \text { holds } .
$$

2 If $f$ is $C^{2}$-smooth and satisfies the estimate

$$
\left\|\nabla^{2} f(y)-\nabla^{2} f(x)\right\|_{\mathrm{op}} \leq M\|y-x\| \quad \text { for all } x, y \in U
$$

then the inequality

$$
\left|f(y)-Q_{x}(y)\right| \leq \frac{M}{6}\|y-x\|^{3}, \quad \text { holds for all } x, y \in U .
$$

## Lipschitz Constants and the Mean Value Theorem

Let $U \subset \mathbf{E}$ be open and $f: U \rightarrow \mathbf{R}$ be $\mathcal{C}^{1}$ on $U$.
Given $x, y \in U$ with $x \neq y$, set $\varphi(t):=f(x+t(y-x))$.
As we have seen $\varphi^{\prime}(t)=\langle\nabla f(x+t(y-x)),(y-x)\rangle$. Hence, by the 1-dimensional mean value theorem (MVT), there exists $\bar{t} \in(0,1)$ such that
$f(y)-f(x)=\varphi(1)-\varphi(0)=\varphi^{\prime}(\bar{t})=\langle\nabla f(x+\bar{t}(y-x)),(y-x)\rangle$.
Consequently, given $z \in U$ and $\epsilon>0$ such that $z+\epsilon \mathbb{B} \subset U$,

$$
|f(y)-f(x)| \leq L\|y-x\| \quad \forall x, y \in B_{\epsilon}(z),
$$

where

$$
L:=\max \{\|\nabla f(v)\| \mid v \in z+\in \mathbb{B}\},
$$

and $\mathbb{B}:=\{x \mid\|x\| \leq 1\}$ is the closed unit ball.
That is, $f$ is locally Lipschitz continuous on $U$ with the local Lipschitz constants given by the gradient. Moreover, if $\mathrm{cl} U$ is compact with $\nabla f$ continuous there, then $L$ an be chosen uniformly for all of $\mathrm{cl} U$.

## Lipschitz Constants and the Mean Value Theorem

Let $U \subset \mathbf{E}$ be open and $F: U \rightarrow \mathbf{R}^{m}$ be $\mathcal{C}^{1}$ on $U$ with component functions $F_{i}$.

Although, there is no MVT for $F$, we do have

$$
F(y)-F(x)=\int_{0}^{1} \nabla F(x+t(y-x))(y-x) d t=\left(\begin{array}{c}
\int_{0}^{1}\left\langle\nabla F_{1}(x+t(y-x)),(y-x)\right\rangle d t \\
\vdots \\
\int_{0}^{1}\left\langle\nabla F_{m}(x+t(y-x)),(y-x)\right\rangle d t
\end{array}\right) .
$$

Hence, given $z \in U$ and $\epsilon>0$ such that $B_{\epsilon}(z) \subset U$,

$$
\|F(y)-F(x)\| \leq L\|y-x\| \quad \forall x, y \in B_{\epsilon}(z)
$$

where

$$
L:=\max \left\{\|\nabla F(v)\|_{o p} \mid v \in z+\epsilon \mathbb{B}\right\}
$$

Again, compactness allows us to choose $L$ uniformly on $\operatorname{cl} U$.

## First-Order Optimality Conditions

Let $f: \mathbf{E} \rightarrow \mathbf{R}$, the directional derivative of $f$ at $x$ in the direction $d$ is given by

$$
f^{\prime}(x ; d):=\lim _{t \downarrow 0} \frac{f(x+t d)-f(x)}{t} .
$$

If $f$ is differentiable at $x$, then $f^{\prime}(x ; d)=\langle\nabla f(x), d\rangle$.

## Theorem (First-order necessary conditions)

Suppose that $x$ is a local minimizer of a function $f: U \rightarrow \mathbf{R}$. Then $f^{\prime}(x ; d) \geq 0$ whenever $f^{\prime}(x ; d)$ exists. If $f$ is differentiable at $x$, then $\nabla f(x)=0$.

## Second-Order Optimality Conditions

## Theorem (Second-order conditions)

Consider a $\mathcal{C}^{2}$-smooth function $f: U \rightarrow \mathbf{R}$ and fix a point $x \in U$. Then the following are true.

1. (Necessary conditions) If $x \in U$ is a local minimizer of $f$, then

$$
\nabla f(x)=0 \quad \text { and } \quad \nabla^{2} f(x) \succeq 0
$$

2. (Sufficient conditions) If the relations

$$
\nabla f(x)=0 \quad \text { and } \quad \nabla^{2} f(x) \succ 0
$$

hold, then $x$ is a local minimizer of $f$. More precisely, it holds:

$$
\liminf _{y \rightarrow x} \frac{f(y)-f(x)}{\frac{1}{2}\|y-x\|^{2}} \geq \lambda_{n}\left(\nabla^{2} f(x)\right)
$$

## Rates of Convergence

Let $\left\{a_{k}\right\} \in \mathbf{R}_{+}$be such that $a_{k} \rightarrow 0$.
Sublinear rate: We will say that $a_{k}$ converges sublinearly if there exist constants $c, q>0$ satisfying

$$
a_{k} \leq \frac{c}{k^{q}} \quad \text { for all } k
$$

Larger $q$ and smaller $c$ indicates faster rates of convergence. In particular, given a target precision $\varepsilon>0$, we have

$$
a_{k} \leq \varepsilon \quad \forall k \geq\left(\frac{c}{\varepsilon}\right)^{1 / q}
$$

The importance of the value of $c$ should not be discounted; the convergence guarantee depends strongly on this value. In applications, it is usually dimension dependent.

## Rates of Convergence

Linear rate: The sequence $a_{k}$ is said to converge linearly if there exist constants $c>0$ and $q \in(0,1]$ satisfying

$$
a_{k} \leq c \cdot(1-q)^{k} \quad \text { for all } k
$$

In this case, we call $(1-q)$ the linear rate of convergence. Fix a target accuracy $\varepsilon>0$, and let us see how large $k$ needs to be to ensure $a_{k} \leq \varepsilon$. Taking logs we get

$$
c \cdot(1-q)^{k} \leq \varepsilon \quad \Longleftrightarrow \quad k \geq \frac{-1}{\ln (1-q)} \ln \left(\frac{c}{\varepsilon}\right) .
$$

Taking into account the inequality $\ln (1-q) \leq-q$, we deduce that

$$
a_{k} \leq \varepsilon \quad \forall k \geq \frac{1}{q} \ln \left(\frac{c}{\varepsilon}\right)
$$

The dependence on $q$ is strong, while the dependence on $c$ is very weak, since the latter appears inside a log.

## Rates of Convergence

Quadratic rate: The sequence $a_{k}$ is said to converge quadratically if there is a constant $c$ satisfying

$$
a_{k+1} \leq c \cdot a_{k}^{2} \quad \text { for all } k
$$

The recurrence yields

$$
a_{k+1} \leq \frac{1}{c}\left(c a_{0}\right)^{2^{k+1}}
$$

The constant $c$ places conditions on when quadratic convergence begins. In particular, if $c a_{0}<1$, then the inequality $a_{k} \leq \varepsilon$ holds for all $k \geq \log _{2} \ln \left(\frac{1}{c \varepsilon}\right)-\log _{2}\left(\ln \left(\frac{1}{c a_{0}}\right)\right)$. The dependence on $c$ is negligible.

Note: $2^{-k}$ converges linearly while $2^{-2^{k}}$ converges quadratically.

