

# CS 446 / ECE 449 — Homework 5

*yiminc2*

## Instructions.

- Homework is due **Friday, November 14**, at 11:59 AM CST; you have **3** late days in total for **all Homeworks**.
- The template for coding problems are available at this link.
- Everyone must submit individually at gradescope under HW5 and HW5 – Programming Assignment.
- The “written” submission at HW5 **must be typed**, and submitted in any format gradescope accepts (to be safe, submit a PDF). You may use L<sup>A</sup>T<sub>E</sub>X, markdown, google docs, MS word, whatever you like; but it must be typed!
- When submitting at HW5, Gradescope will ask you to **mark out boxes around each of your answers**; please do this precisely!
- Please make sure your NetID is clear and large on the first page of the homework.
- Your solution **must** be written in your own words. Please see the course webpage for full **academic integrity** information. You should cite any external reference you use. **The use of LLM is not allowed.**
- We reserve the right to reduce the auto-graded score for HW5 – Programming Assignment if we detect funny business (e.g., your solution lacks any algorithm and hard-codes answers you obtained from someone else, or simply via trial-and-error with the autograder).
- When submitting to HW5 – Programming Assignment, only upload hw5\_q1.py, hw5\_q3.py. Additional files will be ignored.

# 1. Decision Tree (40 pt)

Consider  $\mathcal{D} = \{(\mathbf{x}^{(i)}, y^{(i)})\}_{i=0}^{N-1}$  where  $N \in \mathbb{N}$  and, for  $i \in \{0, \dots, N-1\}$ ,  $\mathbf{x}^{(i)} \in \mathbb{R}^D$  and  $y^{(i)} \in \{0, \dots, C-1\}$  for  $D, C \in \mathbb{N}$  (note: here, we adopt the convention that  $0 \notin \mathbb{N}$ ). In words, we have a training dataset of  $N$  data points, each with  $D$  *continuous* features and belonging to exactly one of  $C$  classes. We wish to train a decision tree classifier using this dataset.

We briefly discussed the decision trees for general continuous spaces. A continuous-valued decision is represented using the total order  $\leq$  on  $\mathbb{R}$ . Specifically, at each node, we will specify a feature  $j$  ( $j \in \{0, \dots, D-1\}$ ) and also a *threshold*  $T$ , and then create *two descendant nodes* at the current node. For all data points in the current node, those with  $x_j$  at most the threshold will be put into the *left child node*, labeled as  $[x_j \leq T]$ , and those with  $x_j$  strictly greater than the threshold will be put into the *right child node*, labeled as  $[x_j > T]$ .

## [Programming] (16 pt)

In this part of the question, you will implement the training and inference of a decision tree classifier on continuous features. We have provided a template of the code in `hw5_q1.py`. Your task is to fill out all the `TODOs` in the methods of the `DecisionTreeClassifier` class.

We provide below a description of methods related to the training of the decision tree classifier.

- `fit` (no `TODO`): Given a training dataset, build the decision tree classifier. This is done by calling the `_grow_tree` method on the entire training dataset.
- `_grow_tree`: Recursively grow the decision tree at a given node (holding a *subset* of the training dataset). The depth of the node is being tracked to avoid growing too large a tree.
- `_find_best_split`: Given a subset of the training dataset, find the split (feature and threshold) that results in the highest information gain. Here, the candidate thresholds for a feature are all possible values of that feature in the subset.
- `_information_gain`: Compute the information gain of a split (feature and threshold). At a parent node, there are  $N_S$  samples and the class distribution is  $\mathbf{p}$ . A split partitions this node into a left child with  $N_L$  samples and class distribution  $\mathbf{p}_L$  and a right child with  $N_R (= N_S - N_L)$  samples and distribution  $\mathbf{p}_R$ . The information gain is defined as:

$$IG(N_S, \mathbf{p}, N_L, \mathbf{p}_L, N_R, \mathbf{p}_R) = H(\mathbf{p}) - \left( \frac{N_L}{N_S} H(\mathbf{p}_L) + \frac{N_R}{N_S} H(\mathbf{p}_R) \right)$$

where  $H$  is the impurity function.

- `_split`: Perform the split on a subset of the dataset by identifying all the indices corresponding to each of the two parts separated based on the specified feature and threshold.
- `_gini`: Given a subset of the labels, compute the Gini impurity. The lecture discussed the use of sample entropy as the impurity measure. Here, we introduced a different impurity measure: Gini impurity, with the following formula

$$\text{Gini}(\mathbf{p}) = 1 - \sum_{i=0}^{C-1} p_i^2$$

where  $\mathbf{p} \in [0, 1]^C$  holds the class distribution in the given subset of labels.

- `_most_common_label`: Given a subset of the labels, find the most common one. If there are many common labels, return the smallest index.

We provide below a description of methods related to the inference of the decision tree classifier.

- `predict` (no `TODO`): Given query samples, make the prediction using the fitted decision tree classifier. This is done by calling the `_traverse_tree` method on each query sample.
- `_traverse_tree`: Given a query sample, recursively traverse the fitted decision tree based on the split at the current node.

**[Written] (24 pt)**

Answer the following questions:

- (a) An impurity measure must be a concave function for the information gain to be non-negative. In this question, we verify that Gini is a qualified impurity measure and explain why concavity is necessary.
- Prove that the Gini impurity is a strictly concave function on  $\mathbb{R}^C$ . **Hint** Show that the Hessian matrix,  $\nabla^2 \text{Gini}(\mathbf{p})$ , is negative definite. (4 pt)  
(Side note This implies that the Gini impurity function is concave on our domain of interest, which is the set of class distributions.)
  - Prove that information gain is always non-negative when the impurity measure is concave. (4 pt)  
**Hint** Find an identity relating  $\mathbf{p}$ ,  $\mathbf{p}_L$ , and  $\mathbf{p}_R$  using  $N_S, N_L, N_R$ . Then, use Jensen's inequality in an appropriate way.
- (b) As discussed in the lecture, a good impurity measure should be minimal when a node is “pure” (i.e., all data belongs to a single class) and maximal when a node is “maximally impure” (i.e., data is uniformly distributed among all classes). In this question, we verify that the Gini impurity has this desired property.
- Let  $\mathbf{p} = (p_0, \dots, p_{C-1})$  be the distribution of classes in a node, where  $p_i \geq 0$  and  $\sum_{i=0}^{C-1} p_i = 1$ .
- Prove that  $\text{Gini}(\mathbf{p}) = 0$  if and only if the node is pure (i.e.,  $p_k = 1$  for some  $k \in \{0, \dots, C-1\}$  and  $p_j = 0$  for all  $j \neq k$ ). (4 pt)
  - Prove that  $\text{Gini}(\mathbf{p})$  is maximized when the class distribution is uniform (i.e.,  $p_i = 1/C$  for all  $i \in \{0, \dots, C-1\}$ ). What is this maximum value? **Hint** Use the method of Lagrange multipliers. (4 pt)
- (c) In this question, we focus on the computational aspect of the decision tree classifier.
- What is the time complexity of the `_find_best_split` method in big-O notation with respect to  $N_S$  (number of data points in the subset  $S$  at the node),  $D$  (number of features), and  $C$  (number of classes)? Explain your answer. (4 pt)
  - Propose a different implementation that will get a better time complexity. What is the improved time complexity in big-O notation with respect to  $N_S$ ,  $D$ , and  $C$ ? Explain your answer. (4 pt)

**Solution.**

## 2. Ensemble methods (Bagging, AdaBoost) (34 pt)

Answer the following questions:

- (a) Fix a training dataset  $\mathcal{D}$  of size  $N$  for a regression task, a query point  $\mathbf{x}$ , and a training algorithm  $\mathcal{A}$  (i.e., a function which takes a dataset and returns a trained model). We create  $B$  bootstrap samples  $\mathcal{D}_1, \dots, \mathcal{D}_B$  by sampling  $N$  times with replacement from  $\mathcal{D}$ . We then train  $B$  models, let  $h_b = \mathcal{A}(\mathcal{D}_b)$ , i.e.,  $h_b$  is the model trained on  $\mathcal{D}_b$ . The bagging predictor is the average:

$$H_B(\mathbf{x}) = \frac{1}{B} \sum_{b=1}^B h_b(\mathbf{x})$$

Note that the only source of randomness in this setup is the bootstrap sampling process (not the training dataset nor the query point).

Since each learner  $h_b$  is trained on a  $\mathcal{D}_b$  drawn from the same bootstrap distribution from  $\mathcal{D}$ , all  $h_b(\mathbf{x})$  are identically distributed. Let this variance be

$$\sigma^2 = \text{Var}[h_b(\mathbf{x})]$$

On the other hand, as discussed in the lecture, any pair of random variables  $h_i(\mathbf{x})$  and  $h_j(\mathbf{x})$  ( $i \neq j$ ) are not independent because the training sets  $\mathcal{D}_i$  and  $\mathcal{D}_j$  are correlated (both are drawn from  $\mathcal{D}$ ). However, all pairs have the same covariance since they are identically distributed. Let this covariance be

$$C = \text{Cov}[h_i(\mathbf{x}), h_j(\mathbf{x})]$$

- i. Derive a formula for the variance of the bagging predictor,  $\text{Var}[H_B(\mathbf{x})]$ , that is dependent on  $B$ ,  $\sigma^2$ , and  $C$ . (2 pt)
  - ii. What does  $\text{Var}[H_B(\mathbf{x})]$  converge to as  $B \rightarrow \infty$ ? Answer in terms of  $\sigma^2$  and  $C$ . (2 pt)
  - iii. Suppose  $\mathcal{A}$  is unstable, in the sense that  $h_i(\mathbf{x}) \neq h_j(\mathbf{x})$  with non-zero probability. Prove that  $C < \sigma^2$  and conclude that  $\text{Var}[H_B(\mathbf{x})] < \sigma^2$  for any  $B > 1$ . **Hint** Consider  $Y = h_i(\mathbf{x}) - h_j(\mathbf{x})$  and compute  $\text{Var}[Y]$ . What can you say about  $\text{Var}[Y]$  if  $\mathcal{A}$  is unstable? (6 pt)  
Side note This means that even with an unstable training algorithm, we will still get reduced variance with bagging.
- (b) We learned from the lecture that the AdaBoost algorithm constructs an ensemble of binary classifiers (with outputs in  $\{-1, +1\}$ ). In this exercise, we explain the design of the AdaBoost algorithm in greater detail.

The AdaBoost algorithm builds an ensemble classifier  $F_T(\mathbf{x}) = \sum_{t=1}^T \alpha_t h_t(\mathbf{x})$  by adding one learner at a time. At step  $t$ , it defines  $F_t(\mathbf{x}) = F_{t-1}(\mathbf{x}) + \alpha_t h_t(\mathbf{x})$  (with  $F_0 = 0$ ) by finding  $(h_t, \alpha_t)$  that minimizes the exponential loss

$$L_t = \sum_{i=1}^N \exp(-y^{(i)} F_t(\mathbf{x}^{(i)})) = \sum_{i=1}^N \exp(-y^{(i)} F_{t-1}(\mathbf{x}^{(i)})) \cdot \exp(-y^{(i)} \alpha_t h_t(\mathbf{x}^{(i)}))$$

Let  $D_t(i)$  be the normalized sample weights from the previous step, defined as

$$D_t(i) = \frac{\exp(-y^{(i)} F_{t-1}(\mathbf{x}^{(i)}))}{\sum_{j=1}^N \exp(-y^{(j)} F_{t-1}(\mathbf{x}^{(j)}))} = \frac{\exp(-y^{(i)} F_{t-1}(\mathbf{x}^{(i)}))}{L_{t-1}}$$

(For  $t = 1$ ,  $F_0 = 0$ , so  $D_1(i) = 1/N$ .)

Since  $L_t = L_{t-1} \cdot \left[ \sum_{i=1}^N D_t(i) \exp(-y^{(i)} \alpha_t h_t(\mathbf{x}^{(i)})) \right]$  and  $L_{t-1}$  is a fixed constant, minimizing  $L_t$  is equivalent to minimizing the normalized objective

$$J_t = \sum_{i=1}^N D_t(i) \exp(-y^{(i)} \alpha_t h_t(\mathbf{x}^{(i)}))$$

- i. Suppose the weak learner  $h_t$  has been chosen to be the learner that best minimizes the weighted error  $\epsilon_t$  defined by

$$\epsilon_t = \sum_{i=1}^N D_t(i) \mathbb{I}(y^{(i)} \neq h_t(\mathbf{x}^{(i)})) = \sum_{\substack{1 \leq i \leq N \\ y^{(i)} \neq h_t(\mathbf{x}^{(i)})}} D_t(i)$$

Assume that  $\epsilon_t \in (0, 1)$ . (If  $\epsilon_t = 0$ , the learner is perfect and we can stop. If  $\epsilon_t = 1$ , we can flip the learner's signs to achieve  $\epsilon_t = 0$ .)

Prove that the optimal weight  $\alpha_t$  that *minimizes* the objective function  $J_t$  is given by:

$$\alpha_t = \frac{1}{2} \ln \left( \frac{1 - \epsilon_t}{\epsilon_t} \right)$$

(6 pt)

**Hint** Split  $L_t$  into two parts: one for incorrectly classified samples and one for correctly classified samples. Then, try to rewrite it in terms of  $\epsilon_t$ . Finally, take the derivative and set it to 0.

- ii. The AdaBoost algorithm updates the sample weights for the next iteration ( $t + 1$ ) using the rule:

$$D_{t+1}(i) = \frac{1}{Z_t} D_t(i) \exp(-y^{(i)} \alpha_t h_t(\mathbf{x}^{(i)}))$$

where  $Z_t$  is a normalization constant (i.e., so that  $\sum_{i=1}^N D_{t+1}(i) = 1$ ).

- A. Using  $\alpha_t$  from the previous part, show that  $Z_t = J_t$ . Derive a formula for  $Z_t$  that is only dependent on  $\epsilon_t$ . (6 pt)
- B. Derive a formula for  $\exp(-y^{(i)} \alpha_t h_t(\mathbf{x}^{(i)}))$  in the two cases  $y^{(i)} = h_t(\mathbf{x}^{(i)})$  and  $y^{(i)} \neq h_t(\mathbf{x}^{(i)})$  that is only dependent on  $\alpha_t$ . (4 pt)

- iii. The training error of the final ensemble  $H_T(\mathbf{x}) = \text{sign}(F_T(\mathbf{x}))$  is

$$\epsilon_{\text{train}} = \frac{1}{N} \sum_{i=1}^N \mathbb{I}(H_T(\mathbf{x}^{(i)}) \neq y^{(i)})$$

Prove that the training error of AdaBoost is exponentially bounded:

$$\epsilon_{\text{train}} \leq \prod_{t=1}^T Z_t$$

(8 pt)

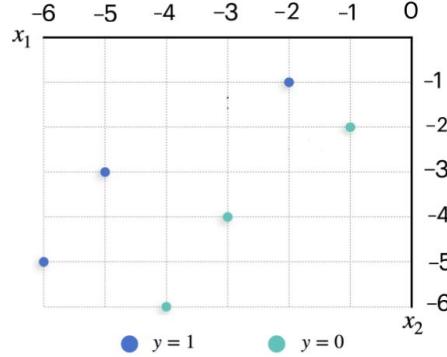
**Hint** Start by showing  $\mathbb{I}(H_T(\mathbf{x}^{(i)}) \neq y^{(i)}) \leq \exp(-y^{(i)} F_T(\mathbf{x}^{(i)}))$ . Then, see if you can relate  $L_T$  to the product of  $Z_t$ .

Side note This implies that, if each weak learner does better than just random chance (i.e.  $\epsilon_t < 0.5$ ), then  $Z_t < 1$ , and  $\epsilon_{\text{train}} \rightarrow 0$  as  $T \rightarrow \infty$ .

**Solution.**

### 3. K-Means (20 pt)

- (a) Suppose we want to perform K-Means clustering on the data samples from the **below** example Fig 3a to cluster them into two groups. Using the algorithm in the slides, perform two runs of the algorithm in the two following settings by showing your work for each step:
- Setting the initial centers to  $\mu_1 = [-1, 0.5]$  and  $\mu_2 = [-5.5, -4.5]$ . (5pt)
  - Setting the initial centers to  $\mu_1 = [0, -4]$  and  $\mu_2 = [-5, 0]$ . (5pt)



- (b) Implement the K-means Algorithm (Lloyd's algorithm) introduced in the lecture, detailed as follows. (10 pt)

---

**Algorithm 1** K-means Clustering (Lloyd's Algorithm)

---

**Require:** Dataset  $\{x^{(i)}\}_{i=1}^N$ , number of clusters  $k$  (i.e., `n_clusters`)

**Ensure:** Cluster centers  $\{\mu_1, \mu_2, \dots, \mu_k\}$  and assignments  $\{C_1, C_2, \dots, C_k\}$

1: **Initialization:** Randomly choose  $k$  examples as initial cluster centers

$$\mu_1^{(0)}, \mu_2^{(0)}, \dots, \mu_k^{(0)}.$$

2: **Iterate** for  $t = 0, 1, 2, \dots$  until convergence:

(1) **Assignment Step:**

Assign each data point  $x^{(i)}$  to its nearest cluster center:

$$C_\ell = \left\{ x^{(i)} : \|x^{(i)} - \mu_\ell^{(t)}\|_2^2 \leq \|x^{(i)} - \mu_j^{(t)}\|_2^2, \forall j \in \{1, \dots, k\} \right\}$$

(2) **Update Step:**

Recompute each cluster centroid as the mean of its assigned points:

$$\mu_\ell^{(t+1)} = \frac{1}{|C_\ell|} \sum_{x^{(i)} \in C_\ell} x^{(i)}, \quad \ell = 1, 2, \dots, k$$

3: **Convergence:** Stop when cluster assignments no longer change.

4: **Output:** Final cluster centers  $\{\mu_\ell\}$  and assignments  $\{C_\ell\}$ .

---

**Solution.**



## 4. Kernels (44 pt)

In this problem, we want to show that new valid kernels can be made by combining existing ones. Suppose  $k_1, k_2$  are valid kernel functions. You need to show:

- (a) i. The set of valid kernels is closed with respect to the scalar product. Formally:  $ck_1(\mathbf{x}^{(1)}, \mathbf{x}^{(2)})$  for any  $c \geq 0$  would be a valid kernel. (4 pt)
- ii. The set of valid kernels is closed with respect to addition. Formally:  $k_1(\mathbf{x}^{(1)}, \mathbf{x}^{(2)}) + k_2(\mathbf{x}^{(1)}, \mathbf{x}^{(2)})$  would be valid kernel. (6 pt)
- iii. The set of valid kernels is closed with respect to multiplication. Formally,  $k_1(\mathbf{x}^{(1)}, \mathbf{x}^{(2)})k_2(\mathbf{x}^{(1)}, \mathbf{x}^{(2)})$  would be valid kernel. (6 pt)
- iv. The set of valid kernels is closed with composition mapping. Formally,  $k_1(f(\mathbf{x}^{(1)}), f(\mathbf{x}^{(2)}))$  would be valid kernel, where  $f : \mathbb{R}^d \rightarrow \mathbb{R}^d$  is a mapping function. (8 pt)
- v. The set of valid kernels is closed with exponentiation. Formally,  $\exp(k_1(\mathbf{x}^{(1)}, \mathbf{x}^{(2)}))$  would be valid kernel. (8 pt)

**Hint:** You can use either of the two following methods for showing that a function  $k : \mathbb{R}^d \times \mathbb{R}^d \rightarrow \mathbb{R}$  is a valid kernel function:

- The matrix

$$K_{ij} = k(\mathbf{x}^{(i)}, \mathbf{x}^{(j)})$$

is symmetric and positive semidefinite for any set of vectors  $\mathbf{x}^{(1)}, \dots, \mathbf{x}^{(n)} \in \mathbb{R}^d$

- $k(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \phi(\mathbf{x}^{(i)})^T \phi(\mathbf{x}^{(j)})$  for some transformation  $\phi$

**Hint:**

$$\exp(x) = \sum_{m=0}^{\infty} \frac{x^m}{m!}.$$

- (b) Recall that the dual problem of a hard-margin SVM is

$$\max_{\boldsymbol{\alpha}} \sum_{i=1}^N \alpha_i - \frac{1}{2} \sum_i^N \sum_j^N \alpha_i \alpha_j y^{(i)} y^{(j)} \mathbf{x}^{(i)\top} \mathbf{x}^{(j)} \quad \text{s.t. } \alpha_i \geq 0, i = 1, \dots, N; \sum_{j=1}^N \alpha_j y^{(j)} = 0$$

Denote the optimal primal solution as  $\mathbf{w}^*$ . We define the domain  $\mathcal{C} = [0, \infty]^N = \{\boldsymbol{\alpha} : \alpha_i \geq 0\}$  for a hard-margin SVM, and  $\mathcal{C} = [0, C]^N = \{\boldsymbol{\alpha} : 0 \leq \alpha_i \leq C\}$  for a soft-margin SVM. We can solve this dual problem by projected gradient descent, which starts from some  $\boldsymbol{\alpha}_0 \in \mathcal{C}$  (e.g.,  $\mathbf{0}$ ) and updates as follows:

$$\boldsymbol{\alpha}_{t+1} = \Pi_{\mathcal{C}} [\boldsymbol{\alpha}_t - \eta \nabla f(\boldsymbol{\alpha}_t)].$$

Here  $\Pi_{\mathcal{C}}[\boldsymbol{\alpha}]$  is the *projection* of  $\boldsymbol{\alpha}$  onto  $\mathcal{C}$ , defined as the closest point to  $\boldsymbol{\alpha}$  in  $\mathcal{C}$ :

$$\Pi_{\mathcal{C}}[\boldsymbol{\alpha}] := \arg \min_{\boldsymbol{\alpha}' \in \mathcal{C}} \|\boldsymbol{\alpha}' - \boldsymbol{\alpha}\|_2.$$

If  $\mathcal{C}$  is convex, in Homework 3, we already prove that

$$\begin{aligned} (\Pi_{[0, \infty)^N}[\boldsymbol{\alpha}])_i &= \max\{\alpha_i, 0\}, \\ (\Pi_{[0, C]^N}[\boldsymbol{\alpha}])_i &= \min\{\max\{0, \alpha_i\}, C\}. \end{aligned}$$

- i. Assume we use radial basis kernel function  $k(\mathbf{x}^{(i)}, \mathbf{x}^{(j)}) = \exp(-\frac{1}{2}\|\mathbf{x}^{(i)} - \mathbf{x}^{(j)}\|_2^2)$ . If the test point  $\mathbf{x}$  is far away from any training point  $\mathbf{x}^{(i)}$ , prove that  $\mathbf{w}^{*\top} \phi(\mathbf{x}) + b \approx b$ . The definition of  $\mathbf{w}^*$  is the same as above. (6 pt)
- ii. On the area  $[-8, 8] \times [-8, 8]$ , plot the contour lines of the following kernel SVMs, trained on the XOR data. Different kernels and XOR data are **provided in** `hw5_utils.py`. Learning rate 0.1 and 10000 steps should be enough. To draw the contour lines, you can use `hw5_utils.svm_contour()`.

- The polynomial kernel with degree 3.
- The RBF kernel with  $\sigma = 1$ .
- The RBF kernel with  $\sigma = 2$ .
- The RBF kernel with  $\sigma = 5$ .

Include these four plots in your **written submission**. (8 pt)

**Solution.**