Appendix: Energy-Based Cross Attention for Bayesian Context Update in Text-to-Image Diffusion Models

390 A **Proof of Theorem 1**

Theorem 1. For the energy functions

$$E(\boldsymbol{Q};\boldsymbol{K}) = \frac{\alpha}{2}\operatorname{diag}(\boldsymbol{K}\boldsymbol{K}^{T}) - \sum_{i=1}^{N}\operatorname{logsumexp}(\boldsymbol{Q}\boldsymbol{k}_{i}^{T},\beta)$$
(17)

392 and

$$E(\boldsymbol{K}) = \log \sum_{i=1}^{N} \exp(\frac{1}{2} \boldsymbol{k}_i \boldsymbol{k}_i^T), \qquad (18)$$

393 the gradient of the log posterior is given by:

$$\nabla_{\boldsymbol{K}} \log p(\boldsymbol{K} \mid \boldsymbol{Q}) = \operatorname{softmax}_{2} \left(\beta \boldsymbol{K} \boldsymbol{Q}^{T} \right) \boldsymbol{Q} - \left(\alpha \boldsymbol{I} + \boldsymbol{D} \left(\operatorname{softmax} \left(\frac{1}{2} \operatorname{diag}(\boldsymbol{K} \boldsymbol{K}^{T}) \right) \right) \right) \boldsymbol{K},$$
(19)

Then, by using the chain rule the update rule of context vectors C is derived as follows:

$$\boldsymbol{C}_{n+1} = \boldsymbol{C}_n + \gamma \bigg(\operatorname{softmax}_2\left(\beta \boldsymbol{K} \boldsymbol{Q}^T\right) \boldsymbol{Q} - \bigg(\alpha \boldsymbol{I} + \boldsymbol{D} \bigg(\operatorname{softmax}\left(\frac{1}{2}\operatorname{diag}(\boldsymbol{K} \boldsymbol{K}^T)\right)\bigg)\bigg) \boldsymbol{K}\bigg) \boldsymbol{W}_K^T,$$
(20)

- where $\gamma > 0$ is a step size, and $D(\cdot)$ is a vector-to-diagonal-matrix operator.
- *Proof.* Based on the Bayes' theorem, the gradient of the log posterior is derived as:

$$\nabla_{\boldsymbol{K}} \log p(\boldsymbol{K} \mid \boldsymbol{Q}) = - \big(\nabla_{\boldsymbol{K}} \operatorname{E}(\boldsymbol{Q}; \boldsymbol{K}) + \nabla_{\boldsymbol{K}} \operatorname{E}(\boldsymbol{K}) \big).$$
(21)

³⁹⁷ First, with definition (17),

$$\nabla_{\boldsymbol{K}} \operatorname{E}(\boldsymbol{Q}; \boldsymbol{K}) = \alpha \boldsymbol{K} - \nabla_{\boldsymbol{K}} \sum_{i=1}^{N} \operatorname{logsumexp}(\boldsymbol{Q} \boldsymbol{k}_{i}^{T}, \beta),$$
(22)

398 where $\forall i \in \{1, \dots, N\}$,

$$\nabla_{\boldsymbol{k}_{i}} \sum_{i=1}^{N} \operatorname{logsumexp}(\boldsymbol{Q}\boldsymbol{k}_{i}^{T}, \beta) = \frac{1}{\beta} \nabla_{\boldsymbol{k}_{i}} \log \sum_{j=1}^{P_{i}^{2}} \exp(\beta \boldsymbol{q}_{j} \boldsymbol{k}_{i}^{T})$$

$$= \sum_{j=1}^{P_{i}^{2}} \frac{\exp(\beta \boldsymbol{q}_{j} \boldsymbol{k}_{i}^{T})}{\sum_{n=1}^{P_{i}^{2}} \exp(\beta \boldsymbol{q}_{n} \boldsymbol{k}_{i}^{T})} \boldsymbol{q}_{j}$$

$$= \operatorname{softmax}(\boldsymbol{Q}\boldsymbol{k}_{i}^{T})^{T} \boldsymbol{Q}.$$
(23)

399 Then, by considering that k_i is a *i*-th row vector of K,

$$\nabla_{\boldsymbol{K}} \sum_{i=1}^{N} \operatorname{logsumexp}(\boldsymbol{Q}\boldsymbol{k}_{i}^{T}, \beta) = \left(\operatorname{softmax}_{1}(\beta \boldsymbol{Q}\boldsymbol{K}^{T})\right)^{T} \boldsymbol{Q}$$

$$= \operatorname{softmax}_{2}(\beta \boldsymbol{K} \boldsymbol{Q}^{T}) \boldsymbol{Q},$$
(24)

- where the last equality holds due to the definition of $softmax_1$ in Section 2.2.
- 401 Second, with definition (18), $\nabla_{\boldsymbol{K}} E(\boldsymbol{K}) = \nabla_{\boldsymbol{K}} \log \sum_{i=1}^{N} \exp(\frac{1}{2} \boldsymbol{k}_i \boldsymbol{k}_i^T)$, where

$$\nabla_{\boldsymbol{k}_{i}} \log \sum_{i=1}^{N} \exp(\frac{1}{2} \boldsymbol{k}_{i} \boldsymbol{k}_{i}^{T}) = \frac{\exp(\frac{1}{2} \boldsymbol{k}_{i} \boldsymbol{k}_{i}^{T})}{\sum_{j=1}^{N} \exp(\frac{1}{2} \boldsymbol{k}_{j} \boldsymbol{k}_{j}^{T})} \boldsymbol{k}_{i}$$

$$= \operatorname{softmax} \left(\frac{1}{2} \operatorname{diag}(\boldsymbol{K}\boldsymbol{K}^{T})\right)_{i} \boldsymbol{k}_{i},$$
(25)

where softmax(\cdot)_{*i*} denotes *i*-th value of a softmax vector. Then,

$$\nabla_{\boldsymbol{K}} \log \sum_{i=1}^{N} \exp(\frac{1}{2} \boldsymbol{k}_{i} \boldsymbol{k}_{i}^{T}) = \boldsymbol{D} \bigg(\operatorname{softmax} \big(\frac{1}{2} \operatorname{diag}(\boldsymbol{K} \boldsymbol{K}^{T}) \big) \bigg) \boldsymbol{K},$$
(26)

where $D(\cdot)$ is a vector-to-diagonal-matrix operator that takes *N*-dimensional softmax vector as an input and returns a $N \times N$ diagonal matrix with softmax values as main diagonal entries. Then, By combining (22), (24) and (26), one can finally obtain:

$$\nabla_{\boldsymbol{K}} \log p(\boldsymbol{K} \mid \boldsymbol{Q}) = \operatorname{softmax}_{2} \left(\beta \boldsymbol{K} \boldsymbol{Q}^{T} \right) \boldsymbol{Q} - \left(\alpha \boldsymbol{I} + \boldsymbol{D} \left(\operatorname{softmax} \left(\frac{1}{2} \operatorname{diag}(\boldsymbol{K} \boldsymbol{K}^{T}) \right) \right) \right) \boldsymbol{K}.$$
(27)

By using the chain rule with $K = CW_K$, the update rule of context vectors C is derived as in (20).

We introduce vector-to-matrix operator $D(\cdot)$ to avoid confusion and fix the typo in the main paper.

409 **B** Pseudo-code for BCU and CACAO

This section provides the description of the pseudocode for the proposed Bayesian Context Update 410 (BCU) and Compositional Averaging of Cross-Attention Output (CACAO). Algorithm 1 outlines 411 the cascaded context propagation across cross-attention layers within the UNet model during the 412 sampling step t. Note that the context is reinitialized at the beginning of each sampling step. On the 413 other hand, Algorithm 2 details the BCU implemented in each cross-attention layer. Remark that 414 D in line 5 denotes vector-to-diagonal-matrix operator. Specifically, the proposed BCU provides a 415 significant computational efficiency by reusing the similarity QK^{T} , which requires computational cost $\mathcal{O}(N^{2})$, to compute $\nabla_{\mathbf{K}} E(\mathbf{Q}; \mathbf{K})$. Consequently, there is only a small amount of additional 416 417 computational overhead associated with the proposed BCU. 418

Algorithm 1 Context cascade at sampling step t	
Require: $Q_t, C_{clip}, UNet$	
1: $C_t \leftarrow C_{clip}$ // Re-initialize	
2: for layer in UNet do	
3: if layer is CrossAttention then	
4: $\boldsymbol{Q}_t, \boldsymbol{C}_t \leftarrow \text{layer}(\boldsymbol{Q}_t, \boldsymbol{C}_t) // \text{Algorithm 2}$	
5: else	
6: $\boldsymbol{Q}_t \leftarrow \text{layer}(\boldsymbol{Q}_t)$	
7: end if	
8: end for	
9: $\boldsymbol{Q}_{t+1} \leftarrow \boldsymbol{Q}_t$	
10: return Q_{t+1}	

Algorithm 2 Bayesian Context Update (BCU)

Require: $Q, C, W_q, W_k, W_v, \alpha, \beta, \gamma_{\text{attn}}, \gamma_{\text{reg}}$ 1: $Q, K, V \leftarrow QW_q, CW_k, CW_v$ 2: $S = QK^T$ 3: $Q \leftarrow \text{softmax}_2(\beta S)V$ 4: $\nabla_K E(Q; K) = \text{softmax}_2(\beta S^T)Q$ 5: $\nabla_K E(K) = -(\alpha I + D(\text{softmax}(\frac{1}{2}\text{diag}(KK^T))))K$ 6: $\Delta C = (\gamma_{\text{attn}} \nabla_K E(Q; K) + \gamma_{\text{reg}} \nabla_K E(K))W_k^T$ 7: $C \leftarrow C + \Delta C$ 8: return Q, C

Algorithm 3 outlines the pseudocode for the CACAO implemented for M given contexts. For the

simplicity, we exclude the BCU from the algorithm. Nontheless, the BCU and the CACAO could be
 leveraged together.

Algorithm 3 Compositional Averaging of Cross-Attention Output (CACAO)

Require: $Q, C = \{C_1, ..., C_M\}, W_q, W_k, W_v, \alpha_s, \beta$ 1: $Q \leftarrow QW_q$ 2: for s in [1, ..., M] do 3: $K_s, V_s \leftarrow C_s W_k, C_s W_v$ 4: $S_s = QK_s^T$ 5: end for 6: $Q \leftarrow \frac{1}{M} \sum_{s=1}^M \alpha_s$ softmax₂ $(\beta S_s) V_s$ 7: return Q

422 C Experimental setups

In this section, we describe detailed experimental setups for three applications including baseline method, hyper-parameter of the proposed method, and dataset if it is the case. Code: https: //github.com/EnergyAttention/Energy-Based-CrossAttention.

426 C.1 Common experimental setup

We mainly leverage pre-trained Stable Diffusion v1-5 (except Table 1: v1-4) which is provided by *diffusers*, a Python library that offers various Stable Diffusion pipelines with pre-trained models. All images are sampled for 50 steps via PNDM sampler [20] using NVIDIA RTX 2080Ti. In every experiment, we set the parameter α in Equation (18) to zero, focusing solely on controlling the values of γ_{attn} and γ_{reg} . BCU is applied to every task, and CACAO is additionally employed in C.4.

Different learning rate for each token It is worth noting that the γ_{attn} and γ_{reg} could be expressed as vectors. In other words, if the context $C \in \mathbb{R}^{N \times d_c}$ is given, γ_{attn} and γ_{reg} are N-dimensional vectors. Hence, we have the flexibility to adjust the learning rate $\gamma_{\{\cdot\}}$, allowing us to increase or decrease the impact of certain tokens based on the user's intent. Unless otherwise noted, γ_{attn} and γ_{reg} is set to a constant for each text token.

⁴³⁷ **Learning rate scheduling** Since the proposed BCU is leveraged for the diffusion model, one can ⁴³⁸ readily introduce scheduling strategies for γ_{attn} and γ_{reg} along the sampling step *t*. We implement ⁴³⁹ multiple variants such as 'constant', 'step', and 'exponential decay' as follows.

[constant]
$$\gamma(t) = \gamma_0$$

[step] $\gamma(t) = \gamma_0 \cdot \text{ReLu}(t - \tau)$ (28)
[exp-decay] $\gamma(t) = \gamma_0 \cdot \lambda^t$

where γ_0 is the initial value, ReLu(x) = 0 if $x \le 0$, otherwise 1, τ denotes the temporal threshold, and λ denotes the decay ratio. Unless stated otherwise, the scheduling strategy is set to the 'constant'.

442 C.2 Multi-concept image generation

We compared the performance of the proposed method with Structured Diffusion [11] which does not require additional training as our method. We leveraged the open-sourced official implementation¹.

For the proposed method, we set the γ_{attn} and γ_{reg} differently for each sample within [1e-2, 1.5e-2, 2e-2]. As shown in the following ablation studies E, large γ_{attn} tends to generate saturated images while large γ_{reg} results in mixed/vanished contents.

We found that using different learning rates for each context token is useful for multi-concept generation, especially when a single concept tends to dominate with a constant learning rate. For example, given the main prompt "A cat wearing a shirt", we set the γ_{attn} for the "shirt" to 3e-2, while γ_{attn} is set to 1.5e-2 for other tokens. We have observed that doubling the γ_{attn} for a text token to be emphasized is sufficient to achieve balanced multi-concept image generation for most cases.

454 C.3 Text-guided image inpainting

Additionally, we conducted a performance comparison between our proposed method and two alternative approaches: (a) Stable Inpaint², which fine-tunes the weights of Stable Diffusion through inpainting training, and (b) Stable Repaint³, which leverages the work of Lugmayr et al. [22] on the latent space of Stable Diffusion for the inpainting task. In the case of Stable Repaint, the mask is downsized and transferred into the latent space. We applied the Bayesian Context Update (BCU) technique to both methods, resulting in improved results compared to their respective baselines.

Masked BCU. To further enhance the performance for the inpainting task, we introduce the concept of masked Bayesian Context Update (masked BCU). Specifically, let $M \in \mathbb{R}^{P_l^2 \times P_l^2}$ represent a diagonal matrix where the main diagonal values are derived from the downsampled inpainting mask for the *l*-th cross-attention layer, with an output spatial size of P_l^2 . In Equation (29), we modify the attention term (12) by incorporating the downsampled mask, effectively covering the query matrix as follows:

$$\boldsymbol{C}_{n+1} = \boldsymbol{C}_n + \gamma \left(\operatorname{softmax}_2\left(\beta \boldsymbol{K} \boldsymbol{Q}^T\right) \boldsymbol{M} \boldsymbol{Q} - \left(\alpha \boldsymbol{I} + \boldsymbol{D} \left(\operatorname{softmax}\left(\frac{1}{2}\operatorname{diag}(\boldsymbol{K} \boldsymbol{K}^T)\right)\right)\right) \boldsymbol{K}\right) \boldsymbol{W}_K^T.$$
(29)

As evident in Equation (12), the attention term updates the context vectors, aligning k_i towards $q_j, j = 1, ..., P_l^2$, while considering the alignment strength between each q_j and k_i . However, in the inpainting task, we have prior knowledge that the context vectors should be most aligned with the semantically relevant masked regions. Therefore, we mask out unrelated background spatial representations, allowing for the context vectors to be updated with a specific focus on the masked regions. This approach facilitates the incorporation of semantic information encoded by k_i specifically into the spatial mask regions.

In our proposed method, we set different values for γ_{attn} and γ_{reg} for each sample, selected from the set [1e-2, 1.5e-2, 2e-2, 2.5e-2], to account for variations in the input samples.

476 C.4 Image editing via compositional generation

We present empirical evidence demonstrating the effectiveness of our energy-based framework for 477 compositional synthetic and real-image editing. The Bayesian Context Update (BCU) technique can 478 be readily applied to both the main context vector (C_1 in Section 3.2, s = 1) and editorial context 479 vectors ($C_{s>1}$). Each BCU operation influences the attention maps used in Compositional Averaging 480 of Cross-Attention Output (CACAO), enhancing the conveyance of semantic information associated 481 with each context. Note that α_s in (16) represents the degree of influence of the s-th concept in the 482 composition. In practice, we fix $\alpha_1 = 1$ for the main context, while $\alpha_{s>1}$ is tuned within the range of 483 (0.5, 1.0).484

¹https://github.com/weixi-feng/Structured-Diffusion-Guidance

²https://huggingface.co/runwayml/stable-diffusion-inpainting

³https://github.com/huggingface/diffusers/tree/main/examples/community# stable-diffusion-repaint

Let $\gamma_{attn,s}$ and $\gamma_{reg,s}$ denote the step sizes for BCU of the *s*-th context vector. If the editing process involves changing the identity of the original image (e.g., transforming a "cat" into a "dog"), we set both $\gamma_{attn,1}$ and $\gamma_{reg,1}$ to zero. Otherwise, if the editing maintains the original identity, we choose values for $\gamma_{attn,1}$ and $\gamma_{reg,1}$ from the range of (5e-4, 1e-3), similar to $\gamma_{attn,(s>1)}$ and $\gamma_{reg,(s>1)}$. All hyperparameters, including α_s and γ_s , are fixed during the quantitative evaluation process (more details in Section D and Table 2).

To ensure consistent results, we maintained a fixed random seed for both real and synthetic image editing. For real image editing, we employed null-text pivotal inversion [24] to obtain the initial noise vector.

During the reverse diffusion process in Sections C.2 and C.3, we kept γ fixed as a constant value. 494 However, for compositional generation, we utilized step scheduling (Equation 28) for γ_s and α_s . 495 After converting the initial noise vector for real images or using a fixed random seed for synthetic 496 images, BCU and CACAO are applied after a threshold time $\tau_s > 0$ for the s-th editorial context. 497 This scheduling strategy helps to preserve the overall structure of generated images during the 498 editing process. In our observations, a value of $\tau_s \in [10, 25]$ generally produces satisfactory results, 499 considering a total number of reverse steps set to 50. However, one can increase or decrease τ_s for 500 more aggressive or conservative editing, respectively. 501

The exemplary real images presented in Figures 5 and 6 of the main paper were sampled from datasets such as FFHQ [16], AFHQ [5], and ImageNet [7]. For a detailed quantitative analysis, please refer to Section D.

505 D Quantitative Comparison

In this section, we conducted a comparative analysis of the proposed framework against several state-of-the-art diffusion-based image editing methods [23, 12, 24, 27], following the experimental setup of [27]. To ensure a fair comparison, all methods utilize the pre-trained Stable Diffusion v1-4, employ the PNDM sampler with an equal number of sampling steps, and adopt the same classifier-free guidance scale.

511 D.1 Baseline Methods

In addition to the Plug-and-Play method discussed in the main paper, we include the following baselines for comprehensive quantitative comparison:

SDEdit [23] + word swap. This method introduces the Gaussian noise of an intermediate timestep and progressively denoises images using a new textual prompt, where the source word (e.g., Cat) is replaced with the target word (e.g., Dog).

Prompt-to-prompt (P2P) [12]. P2P edits generated images by leveraging explicit attention maps from a source image. The source attention maps M_t are used to inject, re-weight, or override the target maps based on the desired editing operation. These original maps act as hard constraints for the edited images.

DDIM + word swap [24]. This method applies null-text inversion to real input images, achieving high-fidelity reconstruction. DDIM sampling is then performed using inverted noise vectors and an edited prompt generated by swapping the source word with the target.

pix2pix-zero [27]. pix2pix-zero first derives a text embedding direction vector $\triangle c_{\text{edit}}$ from the source to the target by using a large bank of diverse sentences generated from a state-of-the-art sentence generator, such as GPT-3 [2]. Inverted noise vectors are denoised with the edited text embedding, $c + \triangle c_{\text{edit}}$, and cross-attention guidance to preserve consensus.

528 D.2 Dataset

For our quantitative evaluations, we focus on three image-to-image translation tasks: (1) translating cats to dogs (cat \rightarrow dog), (2) translating horses to zebras (horse \rightarrow zebra), and (3) adding glasses to cat input images (cat \rightarrow cat with glasses). Following the data collection protocol of [27], we retrieve 250 relevant cat images and 213 horse images from the LAION 5B dataset [33] using CLIP embeddings of the source text description. We select images with a high CLIP similarity to the source word for each task.

535 D.3 Metrics

Motivated by [38, 27], we measure CLIP Accuracy and DINO-ViT structure distance. Specifically, (a) CLIP Acc represents whether the targeted semantic contents are well reflected in the generated images. It calculates the percentage of instances where the edited image has a higher similarity to the target text, as measured by CLIP, than to the original source text [27]. On the other hand, (b) structure distance [38, 37] measures whether the overall structure of the input image is well preserved. It is defined as the difference in self-similarity of the keys extracted from the attention module at the deepest DINO-ViT [3] layer.

543 D.4 Details

The main context vector C_{main} is encoded given a main prompt automatically generated by BLIP [19]. In addition, the editorial context vectors C_{src} and C_{tgt} are encoded given the text descriptions of the source and target concept, i.e. source and target prompt. For example, for a cat \rightarrow dog task (cat \rightarrow cat w/ glasses), the source prompt is "cat" ("cat wearing glasses"), and the target prompt is "dog" ("without glasses"). Then we apply BCU and CACAO based on the obtained context vectors. Please refer to Table 2 for the hyperparameter configurations.

550 D.5 Results

Table 1 shows that the proposed energy-based framework gets a high CLIP-Acc while having low Structure Dist. It implies that the proposed framework can perform the best edit while still retaining the structure of the original input image. This is a remarkable result considering that the proposed framework is not specially designed for the real-image editing task. Moreover, the proposed framework does not rely on the large bank of prompts and editing vector $\triangle c_{\text{edit}}$ [27] which can be easily incorporated into our method.

⁵⁵⁷ While DDIM + word swap records remarkably high CLIP-Acc in horse \rightarrow zebra task, Figure 7 and ⁵⁵⁸ 12 show that such improvements are based on unintended changes in the overall structure. Table 2 ⁵⁵⁹ summarizes the hyperparameter settings for each task. Examples of results are presented in Figure 13 ⁵⁶⁰ and 12.

Table 1: Comparison to state-of-the-art diffusion-based editing methods. Dist for DINO-ViT Structure distance. Baseline results are from [27].

Method	(a) $Cat \rightarrow Dog$		(b) Horse \rightarrow Zebra		(c) Cat \rightarrow Cat w/ glasses	
	CLIP Acc (\uparrow)	Dist (\downarrow)	CLIP Acc (\uparrow)	$\text{Dist}\left(\downarrow\right)$	CLIP Acc (\uparrow)	Dist (↓)
SDEdit [23] + word swap	71.2%	0.081	92.2%	0.105	34.0%	0.082
DDIM + word swap	72.0%	0.087	94.0 %	0.123	37.6%	0.085
prompt-to-prompt [12]	66.0%	0.080	18.4%	0.095	69.6%	0.081
pix2pix-zero [27]	92.4%	0.044	75.2%	0.066	71.2%	0.028
Stable Diffusion + ours	93.7 %	0.040	90.4%	0.061	81.1 %	0.052

Table 2: Hyperparameter configurations for each editing task. Each task index comes from Table 1. $\gamma_{attn,main} = 0$ and $\gamma_{reg,main} = 0$ as mentioned in section C.4. Note that $\alpha_{src} < 0$ for the concept negation (related ablation study in Figure 9). τ_s denotes the warm-up period for step scheduling in (28) and Section C.4.

Task	α_{src}	α_{tgt}	$\gamma_{\cdot,main}$	$\gamma_{attn,src}$	$\gamma_{reg,src}$	$\gamma_{attn,tgt}$	$\gamma_{reg,tgt}$	$ au_s$
(a)	-0.65	0.75	0	5e-4	5e-4	6e-4	6e-4	25
(b)	-0.5	0.6	0	4e-4	4e-4	5e-4	5e-4	15
(c)	-0.6	0.7	0	1e-3	1e-3	1e-3	1e-3	17



Figure 7: Image editing comparison with DDIM-inversion. Generated samples by DDIM-inversion with word swap readily deviate the original data contents, while the proposed method avoids undesired changes.

561 E Ablation study and more results



Figure 8: Ablation results for γ_{attn} and γ_{reg} . All samples are generated from the same random noise.

"A teddy bear"



Figure 9: Ablation study results. The first row shows multi-concept generation examples with varying γ_{attn} and γ_{reg} , while the second row shows real image editing examples with varying the usage of BCU and CACAO. The last row shows the effect of negative prompt for the image editing application.

Attention and regularization terms. To access the degree of performance improvement attained 562 by the proposed BCU, we conducted an ablation study for the attention and the regularization terms 563 by regulating γ_{attn} and γ_{reg} for the text-guided image inpainting (Figure 8) and the multi-concept 564 image generation (Figure 9). From the Figure 8, we can observe that the desired content is generated 565 when proper range of γ_{attn} and γ_{reg} are given. Specifically, once γ_{reg} is set to a valid value, the BCU 566 consistently generate a "teddy bear" with various γ_{attn} , otherwise it generates background or 567 imperfect objects. This result emphasizes the role of the introduced prior energy $E(\mathbf{K})$. Furthermore, 568 the γ_{attn} also affects to the context alignment of the generated sample (for instance $\gamma_{attn} = 0.025$ 569 and $\gamma_{req} = 0.02$), which highlights the importance of the introduced conditional energy function 570 $E(Q; \vec{K})$. The same evidences could be found in the first row in Figure 9 which are the multi-concept 571 image generation examples. 572

Synergy between BCU and CACAO. While both BCU and CACAO are designed from the common 573 energy-based perspective, each operation is originated from different energy functions $E(\mathbf{K}; \mathbf{Q})$ and 574 $\hat{E}(Q; \{K_s\}_{s=1}^M)$, respectively. This fact suggests the synergistic energy minimization by combining 575 the BCU and CACAO, which could further improve the text-conditional image generation. To inves-576 577 tigate this further, we conducted an ablation study using a real image editing application. Specifically, we compared the editing performance when solely utilizing CACAO and when combining BCU with 578 CACAO. The second row in Figure 9 is the result of the ablation study that shows fully-compatibility 579 of the BCU and CACAO. Importantly, the incorporation of the BCU improves the quality of the 580 generated images. While the CACAO alone effectively captures the context of the given editing 581 concept, the addition of BCU enhances the fine-grained details in the generated outputs. 582

Importance of concept negation. Remark that a negative α_s in (16) denotes the negation of given 583 editing prompt. We empirically observed that the concept negation may significantly contribute to 584 the performance of compositional generation. Specifically, for the image-to-image translation task in 585 Table 1, we apply both positive and negative guidance with the target (e.g. Dog) and source (e.g. Cat) 586 concepts, respectively, following the degree of guidance denoted in Table 2. The third row in Figure 9 587 shows the impacts of source concept negation in the image-to-image translation task. While the 588 positive guidance alone may fail to remove the source-concept-related features, e.g. eyes of the Cat, 589 the negative guidance removes such conflicting existing attributes. This implies that the proposed 590 framework enables useful arithmetic of multiple concepts for both real and synthetic image editing. 591

Prior energy and α . While $\frac{\alpha}{2}$ diag(KK^T) in (7) penalizes norm of each context vectors uniformly, the proposed prior energy function E(K) adaptively regularizes the smooth maximum of $||k_i||$. Intuitively, adaptive penalization prevents the excessive suppression of context vectors, potentially resulting in images that are more semantically aligned with a given context. To demonstrate the effectiveness of adaptive penalization in the prior energy function, we conducted a multi-concept image generation task with varying α in (20) from 0 to 1, while fixing other hyperparameters. Figure 10 illustrates the gradual disappearance of salient contextual elements in the generated images depending on the change of α . Specifically, the crown is the first to diminish, followed by subsequent context elements, with the lion being the last to vanish with $\alpha = 1$. This result highlights the validity of the adaptive penalization for the context vectors which stems from the prior energy function.



Figure 10: Generated samples with varying α values. As α increases, the generated images progressively deviate from the intended context, "A lion and a crown".



Figure 11: Further results for multi-concept image generation. Best views are displayed.



Figure 12: Further results for real image editing: horse to zebra.



Figure 13: Further results for real image editing: cat to dog.



Figure 14: Further results for image editing with varying text prompts. Best views are displayed.

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