CFDiffusion: Controllable Foreground Relighting in Image Compositing via Diffusion Model

Anonymous Author(s)



Composite image

Image harmonization

Shadow generation

Figure 1: In this illustration, we demonstrate the specific effects of image harmonization and shadow generation. The areas enclosed by the red rectangles represent our foreground objects. From the composite image to the ground truth (GT), we integrate these two operations into a unified model: adjusting the appearance of the foreground object to harmonize it with the background and generating reasonable shadows for the foreground object.

ABSTRACT

Inserting foreground objects into specific background scenes and eliminating the illumination inconsistency (eg., color, brightness) between them is an important and challenging task. It typically involves multiple processing tasks, such as image harmonization and shadow generation. In these two domains, there are already many mature solutions, but they often only focus on one of the tasks. Recently, some image composition methods have utilized diffusion models to address both of these issues simultaneously, but they cannot guarantee the complete reconstruction of foreground content. In this work, we propose CFDiffusion, which can simultaneously handle image harmonization and shadow generation. We first employ a shadow mask predictor to estimate the shadow mask of the foreground object. Next, we design a harmonization-shadow generator based on a diffusion model to harmonize the foreground and generate shadows concurrently. Additionally, we propose a foreground content enhancement module to ensure the complete preservation of foreground content at the insertion location, and we also develpp an adaptive encoder to guide the harmonization process in the foreground area. The experimental results on the iHarmony4 dataset and our IH-SG dataset demonstrate the superiority of our CFDiffusion approach.

KEYWORDS

CFDiffusion, image composition, object relighting, image harmonization, shadow generation

1 INTRODUCTION

Image composition stands as a fundamental task in computer vision and augmented reality, aimed at seamlessly integrating objects from one image into another to craft a convincingly realistic composite image. Merely inserting the foreground object into a background image without careful consideration results in noticeable discrepancies between the foreground and background, such as differences in color, brightness, and shadows. Based on this, we can decompose the image compositing process into multiple subtasks, each addressing specific issues: image harmonization [2, 5, 6, 9, 19, 23, 48, 51] and shadow generation [17, 28, 44, 55].

Image harmonization aims to adjust the compatibility between the foreground and background in terms of color and brightness, while shadow generation ensures that the inserted foreground objects cast realistic and reasonable shadows. Various practical methods exist for both of these subtasks. However, employing multiple models to address each subproblem individually is both cumbersome and impractical. What we need is a unified network that can address both of these issues simultaneously, achieving excellent results for each. Figure 1 illustrates the problems we need to deal with and the results we should achieve

In recent years, generative models such as GANs [3, 18, 20, 43] and diffusion models [1, 16, 30, 34, 36, 39, 40] have demonstrated significant potential in image composition. Particularly, diffusion models have surpassed various preceding methods in image editing [1, 22, 34] and other applications [15, 33, 37]. Conditional diffusion model aims to generate images under the guidance of conditional information, such as text or semantic masks. Among them, Stable

117

118

151

152

153

154

155

156

157

158

159

160

161

162

163

164

165

166

167

168

169

170

171

172

173

174

Diffusion (SD) [39] stands out as one of the most popular models, successfully integrating text from CLIP [38] into latent diffusion.

119 Some existing works have introduced the diffusion model into related research domains, particularly in image editing and image 120 composition. For example, SDEdit [34] composites images by adding 121 noise to the input image and then iteratively denoises it through sto-123 chastic differential equations. However, these methods lacks proper 124 and sufficient guidance during the denoising process, resulting in 125 the final image lacking sufficient content fedelity. Besides, most 126 diffusion models for image editing focus on manipulating images using text input, which is inappropriate for image composition. 127

Recently, some image compositing methods [46, 52] have at-128 tempted to address all issues within a unified model, which can 129 significantly simplify both model size and complexity. For exam-130 ple, ObjectStitch [46] utilizes a bounding box that encompasses 131 the foreground object to specify the region for foreground object 132 insertion and shadow generation, then processes the foreground 133 within this designated area. Given the recent successful applications 134 135 of diffusion models in image processing, these methods typically rely on pretrained diffusion models. However, in practice, these 136 137 methods result in uncontrollable adjustments to the foreground in 138 terms of both position and content texture, raising concerns about 139 preserving the fidelity and credibility of the foreground object.

In this paper, our aim is to address the issue of inadequate fore-140 ground fidelity observed in previous discussions on image com-141 142 position. We introduce a method called CFDiffusion that concurrently handles image harmonization and shadow generation tasks. 143 Building upon stable diffusion, we introduce a foreground content 144 enhancement module (FCEM), which utilizes a foreground content 145 encoder to extract foreground content information, thus guiding 146 the reconstruction of foreground content. Furthermore, we equip 147 148 SD with a lightweight adaptive encoder designed to extract cru-149 cial conditional information from the composite image, such as background style and color, to guide the denoising process of SD. 150

To validate the effectiveness of our approach, we compare it with state-of-the-art methods and conduct experiments on benchmark datasets such as iHarmony4 [7] and our proposed IH-SG dataset. The experimental results demonstrate that our method achieves more realistic image harmonization and produces shadows that are both genuine and believable.

Our contributions can be summarized as follows:

- We introduce a novel image composition method called CFDiffusion. This method simultaneously handles image harmonization and shadow generation tasks for foreground objects with masked insertion points.
- We design a foreground content enhancement module to fully reconstruct the foreground content and texture details.

Extensive experiments conducted on both public datasets and our newly created dataset IH-SG validate the effectiveness of our proposed method.

2 RELATED WORK

2.1 Image Harmonization

As a subtask of image compositing, the objective of image harmonization is to integrate objects from a given foreground image into a background image to create a cohesive composite image. This 175

176

177

178

179

180

181

182

183

184

185

186

187

188

189

190

191

192

193

194

195

196

197

198

199

200

201

202

203

204

205

206

207

208

209

210

211

212

213

214

215

216

217

218

219

220

221

222

223

224

225

226

227

228

229

230

231

232

process involves adjusting the color and lighting information of the foreground object to ensure its compatibility with the background of the composite image.

Traditional methods [26, 47] rely on adjusting the appearance of the foreground to match the color statistics of the background, typically focusing on obtaining color statistics and then transferring this information between the foreground and background. These methods are fast and straightforward but often struggle with complex scenes and produce artifacts because the realism of the image is often not well captured by these statistics.

Particularly, with the release of the first large-scale image harmonization dataset, iHarmony4 [7], supervised image harmonization methods [4, 8, 10, 11, 13, 27] have garnered increasing attention. For instance, [11] employed attention blocks to compute non-local information for foreground adjustment. SSAM [8] integrates them using a dual-path attention model, focusing on the relationship between spliced and unspliced regions. DoveNet [7] treats image harmonization as a domain translation task. CDT-Net [6] combines pixel-to-pixel and RGB-to-RGB transformations for high-resolution image harmonization. [27] introduced the concept of style from the background image, treating the harmonization task as a style transfer problem. They proposed a novel Region-Aware Instance Normalization (RAIN) method, which extracts style information solely from the background features and applies it to the foreground of the image harmonization task. However, when the task extends to shadow generation, these methods do not scale well to handle both tasks simultaneously.

2.2 Shadow Generation

Previous work on shadow generation can be categorized into two main approaches: rendering-based methods and image-to-image translation methods.

Rendering-based methods [21, 24] relies on a clear understanding of lighting, reflectance, material properties, and scene geometry to generate shadows for inserted virtual objects using rendering techniques. However, such knowledge is often unavailable or impractical for applications in real-world scenarios. Image-to-image translation methods predominantly employ deep learning techniques, characterized by encoder-decoder architectures. By training on paired images, including images with shadows and those without, these methods directly learn the mapping from shadow-free images to shadowed images from the input data. Importantly, this approach typically eliminates the need for explicit knowledge about lighting, reflectance, material properties, and scene geometry. The ARShadowGAN [28] model introduces an attention-guided network capable of directly modeling the mapping relationship between the shadows of foreground objects and their corresponding real environments, accompanied by the release of the Shadow-AR dataset. SGRNet [17] promotes comprehensive information interaction between the foreground and background. It initially predicts a mask for shadow regions and subsequently forecasts shadow parameters to fill these regions. Additionally, a new shadow training dataset, DESOBA, is introduced. ShadowGAN [55] combines both global conditional discriminators and local conditional discriminators to generate shadows for inserted 3D foreground objects without relying on background lighting information.Shadow generation is



Figure 2: The display of the data pairs of the dataset we created, from left to right: ground truth, composite image, foreground object mask, shadow mask of foreground object, background object mask, shadow mask of the background object. We captured foreground and background separately under different lighting environments, paying attention to collecting various scenes, ground surfaces, and shadow casting situations.

associated with the foreground objects, but it targets different areas than image harmonization. We aim to combine both aspects using a single network framework.

3 PROPOSED METHOD

Given a composite image I_c , a binary mask M_{bs} representing the background object-shadow pair, and a binary mask M_f indicating the foreground object, our goal is to obtain an image \tilde{I} that harmonizes the foreground object and produces reliable shadows under background illumination conditions.

As illustrated in Figure 4, our method consists of two stages: foreground object shadow mask prediction stage and shadow-harmonization generation stage. It mainly comprises four components: foreground object shadow generator G_{fs} , foreground content encoder $E(\cdot)$, adaptive encoder E_a , foreground content enhancement module *FCEM*, and shadow-harmonization generator $G(\cdot)$ based on a stable diffusion model.

The network workflow is as follows: Firstly, we use the background object-shadow data pair as reference to predict the shadow mask of the foreground object, identifying the approximate location for shadow generation. Then, we input the synthesized image into the shadow-harmonization generator to produce the final result. Simultaneously, we use the foreground content encoder to extract the foreground content embedding, inputting it into the foreground content enhancement module to constrain and complete foreground texture details. The adaptive encoder transfers the background style to the foreground region, providing additional generation guidance for the harmonization-shadow generator.

3.1 Harmony-Shadow Generator

Recently, diffusion models have shown remarkable performance in many fields: image generation [16, 45], text-to-image generation [39], image translation [25], image inpainting [32, 41], and image editing [12, 34]. The backbone of our Harmony-Shadow Generator is built upon a Stable Diffusion (SD) [39] model.

SD is a latent diffusion model that undergoes a two-stage pretraining process, involving an autoencoder and a denoising U-Net [16]. In the first stage, the SD model trains an autoencoder: the encoder \mathcal{E} converts the images I into a latent representation $z'_0 = \mathcal{E}(I)$, and then the decoder \mathcal{D} reconstructs the images, resulting in



Figure 3: Overview of Cross-Attention Integration (Cross-Attention Integration) layer [17]. The g, f, h, v shown in the figure represent 1×1 convolution, \otimes represents matrix multiplication.

 $\tilde{I} = \mathcal{D}(z'_0)$. In the second stage, the autoencoder's parameters are fixed, and SD introduces noise to the latent space representation z'_0 over T steps to generate z'_t . This process involves the creation of a denoising U-Net ϵ_{θ} , which is trained using a latent denoising loss

$$\mathcal{L}_{LDM}: = \mathbb{E}_{z'_0, y, \epsilon} \mathcal{N}_{(0,1), t} [\parallel \epsilon - \epsilon_{\theta_1}(z'_t, t, \tau_{\theta_2}(y)) \parallel_2^2], \qquad (1)$$

Here ϵ is the noise added to the latent space feature z'_0 at each noise step, ϵ_{θ_1} is the denoising U-Net that predicts the noise ϵ at the current step t, and y represents additional conditions (e.g. text, mask, etc.), τ_{θ_2} is instead a domain-specific encoder that projects y to an intermediate representation.

In this work, we add conditional information using an adaptive encoder similar to that of composite images with foreground masks. During the inference process, noise is first added to z'_0 to generate z'_T , and then z'_T is used as z_T , which is the initial input of ϵ_{θ_1} . Then iteratively use ϵ_{θ_1} to estimate the noise at each denoising step t, thereby gradually refining the latent map z_T , and ultimately become a clean latent feature z_0 . Finally, the clean latent features z_0 are fed to the decoder \mathcal{D} to generate images.

: Foreground content

enhancement module

 E_a : Adaptive encoder

FCEM



FCEN

Diffusion U-Net Mode

Figure 4: Overview of our CFDiffusion. It consists of two stages: foreground object shadow mask prediction stage and shadowharmonization generation stage. It mainly comprises four components: foreground object shadow generator G_{fs} , foreground content encoder $E(\cdot)$, adaptive encoder E_a , foreground content enhancement module *FCEM*, and shadow-harmonization generator $G(\cdot)$. E_f and E_b respectively represent the foreground encoder and background encoder of the foreground object shadow mask generator.

3.2 Foreground shadow mask generator

Inspired by [17], we apply a shadow mask predictor to generate foreground object shadow predictor. First, we predict the foreground shadow mask \widetilde{M}_{fs} through the foreground shadow mask generator G_s . G_s consists of an encoder and a decoder D, and the encoder is divided into a foreground encoder E_f and a background encoder E_b . We believe that the background object-shadow pair contains clues that are beneficial for inferring the foreground shadow area. In order to generate the shadow mask of the foreground object, we take the concatenation of the composite image I_c and the background object-shadow mask M_{bs} as the input of the background encoder E_b , and generate the background feature map X_b . At the same time, the concatenation of the composite image I_c and the foreground object mask M_f is used as the input of the foreground encoder E_f to obtain the foreground feature map X_f . The process is summarized as follows:

$$X_b = E_b \left(M_{bs}, \ I_c \right), \tag{2}$$

$$X_f = E_f \left(M_f, \ I_c \right). \tag{3}$$

Following [17], we use a Cross-Attention Integration (CAI) [17, 49, 50, 54] layer to help the foreground feature map notice the relevant lighting information of the background feature map. As the picture 3 shows, the input of the CAI layer consists of X_f and X_b , which are outputs of foreground encoder E_f and a background encoder E_b , and the output feature map is denoted as X. Then X is fed into the decoder D to obtain the mask of the foreground object shadow. Subsequently, we add it to the foreground object mask M_f to obtain the foreground object-shadow mask \tilde{M}_{fs} , which serves as one of the inputs for the subsequent shadow-harmonization generator. The process is summarized as follows:

$$\widetilde{M}_{fs} = D(X) + M_f. \tag{4}$$

3.3 Foreground Encoder

Following [53], in order to further enhance the detailed texture of the foreground generation, we employ the pre-trained model ViT - L/14 from CLIP [38] as the foreground image encoder E. Initially, we extract the foreground object region from the synthesized image I_c using the foreground object mask M_f , which is then inputted into the foreground image encoder E to extract the local content embedding of the foreground E_l . This process can be represented as follows:

Denoise Step

$$E_l = E(I_f). \tag{5}$$

The intermediate layer of the CLIP encoder outputs 256 patch tokens containing local details. We extract the information of these patch tokens and integrate these foreground content embeddings into the Foreground content enhancement module (FCEM) of the denoising U-Net model to help us control the generation of foreground content details. The specific details of the FCEM module is located in Section 3.4.

3.4 Foreground content enhancement module(FCEM)

Following [53], we utilize a foreground content enhancement module to embed foreground content into the intermediate features of the diffusion model, thereby constraining the stable diffusion model for foreground appearance generation and promoting the composite generation of foreground appearance with high fidelity.

Our foreground content enhancement module is built upon the publicly released v1-4 SD model. To identify foreground regions that need to be constrained, we append the binary foreground object-shadow mask \tilde{M}_{fs} to the model input. To achieve this, in the first convolutional layer of U-Net, we attach two additional input channels to respectively contain the foreground object mask M_{fs}

Anon



Figure 5: Overview of FCEM module. A in the figure is the attention map output by the cross attention part. $F_i \in \mathbb{R}^{h_i \times w_i \times c_i}$ is feature map of i-th transformer block

and the predicted foreground object-shadow mask \widetilde{M}_{fs} . Eventually, the input images are uniformly resized to a resolution of 256×256 .

Denoising U-Net of SD consists of a series of basic blocks, each block includes a residual block and a transformer block. The transformer block consists of a self-attention module, a cross-attention module, and a feedforward network.

As illustrated in Figure 5. We record the features from the i - th transformer block as $F_i \in \mathbb{R}^{h_i \times w_i \times c_i}$, where h_i, w_i, c_i represent its height, width, and channel dimensions respectively. We first use the foreground local feature F_i^l intercepted by the foreground object mask (M_f) resized to h_i, w_i, c_i . The feature map will be flattened to $\overline{F}_i^l \in \mathbb{R}^{N \times c_i}$, and then passed through cross attention together with the foreground local embedding E_l , and then we get an attention map A and refined foreground local feature \overline{F}_i^l to obtain the aligned foreground embedding map \widetilde{E}_l . Further use \widetilde{E}_l to modulate \overline{F}_i^l . \widetilde{E}_l is passed through a convolutional layer of 3×3 to obtain the spatial awareness modulation weight, and the modulation is normalized The transformed \overline{F}_i^l is as follows:

$$\hat{F}_{i}^{l} = norm(\widetilde{F}_{i}^{l}) \bullet conv(\widetilde{E}_{l}).$$
(6)

Finally, after resizing \hat{F}_i^l , it is added to the foreground object region of F_i . The output of the Foreground Content Enhancement Module (FCEM) is then delivered as the enhanced foreground content features \tilde{F}_i to the next residual block.

3.5 Adaptive Encoder

Following [31, 35], we adopt an adaptive encoder, which is a lightweight model that can align the internal knowledge in the SD model
with external control signals. Through this adaptive encoder, we
can achieve rich control effects on the color and structure of the
SD generation results.

Conference acronym 'XX, June 03-05, 2018, Woodstock, NY

The adaptive encoder takes into account encoding additional conditions and provides multi-step guidance for denoising U-Net in the denoising step. Previous adaptive encoder implementations focused more on coarse structures (e.g., sketches, poses, semantic masks) and exploited textual conditions to indicate additional requirements (e.g., style or context). Different from previous work, we abandon the text CLIP model, splice the composite image I_c and the foreground mask M_f , and use a lightweight adaptive encoder to encode while retaining content details and extracting background styles. The structure of the adaptive encoder includes four feature extraction blocks and three DownSample (DS) blocks.

First, the input image will be resized to 64×64 , and we name it F_c^0 . Each feature extraction module (EM) includes a convolutional layer and two residual blocks. The generation process of F_c^i , $i \in 1, 2, 3, 4$ can be expressed as follows:

$$F_c^1 = EM_1\left(F_c^0\right),\tag{7}$$

$$F_c^i = EM_i \left(DS(F_c^{i-1}) \right). \tag{8}$$

The resolutions of F_c^1, F_c^2, F_c^3 , and F_c^4 are 64×64, 32×32, 16×16, 8×8 respectively. Then we use foreground mask M_f to separate the foreground features $F_{c,f}^i$ and background features $F_{c,b}^i$:

$$F_{c\,f}^{i} = Flatten(F_{c}^{i} \circ M_{f}), \tag{9}$$

$$F_{c\,b}^{i} = Flatten(F_{c}^{i}(1-M_{f})).$$
⁽¹⁰⁾

Among them, M_f represents the foreground object mask scaled to the corresponding F_c^i size, \circ represents element-wise product, and $Flatten(\cdot)$ represents expanding a 2D feature map into a 1D feature feature sequence.

We use a transformer layer to extract and transfer the style of the background to the foreground area to achieve harmonious processing of the foreground. In addition, the parts of the background area that are related to the foreground can provide more references when harmonizing the foreground, so they are very important. We will also pay more attention to the areas that are more related to the background and foreground. $F_{c,f}^i$ is used as query, $F_{c,b}^i$ is used as keys/values, and the final background stylized foreground feature $\hat{F}_{c,f}^i$ can be expressed as:

$$\hat{F}_{c,f}^{i} = Transformer(F_{c,f}^{i}, F_{c,b}^{i}, F_{c,b}^{i}).$$
(11)

3.6 Traning Losses and Details

Our total loss function L_{total} consists of the standard noise loss \mathcal{L}_{LDM} of the diffusion model and a reconstruction loss L_{rec} . Therefore, the final loss function of our CFDiffusion is:

$$L_{total} = \lambda_1 L_{rec} + \lambda_2 \mathcal{L}_{LDM} + \lambda_3 L_{fs}, \tag{12}$$

where $\lambda_1, \lambda_2, \lambda_3$ are hyper-parameters which control the influence of terms.

Noise Loss. First, we adopt the standard noise loss of the diffusion model, aiming to reconstruct the image features in the latent space, shown as Equation (13):

$$\mathcal{L}_{LDM} := \mathbb{E}_{z'_0, y, \epsilon \ \mathcal{N}(0,1), t} \left[\parallel \epsilon - \epsilon_{\theta_1}(z'_t, t, \tau_{\theta_2}(y)) \parallel_2^2 \right].$$
(13)

Reconstruction loss. It is a classical L_1 loss between the generator output image \hat{I} and real ground-truth image I, to further constrain



Figure 6: The comparison between our method and two image harmonization methods: DucoNet [48] and DIH-GAN [2], three shadow generation methods [28, 29, 46] on our dataset. It can be clearly seen that our CFDiffusion has achieved the best results in both real shadow generation and image harmonization.

the generated image towards the ground truth, which is expressed as:

$$L_{rec} = \parallel \widetilde{I} - I \parallel_1.$$
⁽¹⁴⁾

MSE Loss. Additionally, we compute the loss for the foreground mask prediction module using the following method:

$$L_{fs} = \| M_{fs} - \tilde{M}_{fs} \|_2^2 .$$
 (15)

4 EXPERIMENTS

To verify the superiority of our proposed CFDiffusion, we compare CFDiffusion with state-of-the-arts on the real-world iHarmony4 [7] dataset and our proposed dataset, and provide assessments both quantitatively and qualitatively.

4.1 Experimental Settings

The proposed method is implemented using PyTorch, which is trained using one NVIDIA RTX 3090 GPU. All images are resized to 256×256 for training and testing. We adopt adam optimizer with the momentum as (0.9,0.999), and the learning rate initialized as 0.00003. Following [42], we use the Kaiming initialization technique [14] to initialize the weights of the proposed model and use a 0.9999 Exponetial Moving Average(EMA) for all our experiments. We used 1000 diffusion steps T and noise schedule β_t linearly increasing from 0.0001 to 0.002 for training, and 25 steps for inference. After a few trials, we set $\lambda_2 = \lambda_3 = 10$, $\lambda_1 = 1$ by observing the grenerated images. The training epoch is set as 1500.

Compared methods. We compared our model with five deep
 learning-based methods from the related fields: two image har monization methods including DucoNet [48], DIH-GAN [2], three
 shadow generation methods include ARShadowGAN [28], SGDif fusion [29], and ObjectStitch [46]. Among them, SGDiffusion is
 the latest method using a diffusion model for shadow generation
 tasks, while ObjectStitch is the newest image composition method

that addresses both image harmonization and shadow generation tasks. DucoNet for image harmonization based on dual color spaces. ARShadowGAN makes full use of the background information to guide the shadow generation of foreground objects. We train and test all these methods based on our dataset. For detailed information about our IH-SG dataset, please refer to the remainder of this section.

In addition, we also tested our image harmonization capabilities on the iHarmony4 dataset, and compared the three image harmonization methods of CDT-Net [6], Harmonizer [23] and DucoNet [48] to further prove the superiority of our CFDiffusion. CDT-Net coherently combines pixel-to-pixel conversion and RGB-to-RGB conversion in an end-to-end network.

Evaluation metrics. We use four metrics to evaluate the image illumination harmonization results, which are Relative Mean Square Error (RMSE), Structural Similarity Index Measure (SSIM), foreground Mean Square Error (fMSE), foreground Structural Similarity Index Measure (fSSIM). Generally, the smaller RMSE and fMSE, and the larger fSSIM and SSIM indicate the better image illumination harmonization results.

Dataset. To better train our model, we have constructed a dataset called IH-SG that can address both of image harmonization and shadow generation concurrently. Each data pair we construct includes: composite images, ground truth images, foreground object masks, foreground object shadow masks, background object masks and background object shadow masks.

In total, our dataset comprises over 1000 outdoor scenes and more than 10000 data pairs. We also captured numerous shadow scenes under complex conditions, such as shadows cast on walls and steps, to enrich the shadow data samples, making our dataset more realistic and diverse. Our data pairs are illustrated in the Figure 2.

4.2 Comparison with State-of-the-Arts

Experiments on our dataset. The quantitative comparison results of different methods on our testing set are summarized in Table 1. Apparently, our method CFDiffusion achieves the better quantitative results than other state-of-the-art methods on all four metrics. On the one hand, this is due to the powerful image generation capability of the diffusion model. On the other hand, our FCEM and adaptive encoder make full use of the foreground and background information, proving the superiority of our CFDiffusion.

Experiments on iHarmony4. The iHarmony4 dataset is one of the most popular large-scale datasets in the field of image harmonization, covering a variety of scenes and foreground objects. Therefore, we compared our method with several image harmonization techniques on the iHarmony4 dataset, and the quantitative comparison results are shown in Table 3. Our proposed approach also achieved the best results.

Experiments on DESOBAv2. Recently, Liu et al. [29] extended the DESOBA [17] dataset to DESOBAv2, which has become the latest shadow generation dataset. To validate the generalization ability of our model, we conducted experiments on the DESOBAv2 dataset by applying slight perturbations to foreground objects, and the results are shown in Figure 13.

4.3 Ablation Study

In order to verify the effectiveness of each component of our method, we conduct ablation studies by modifying the CFDiffusion architecture. Specifically, we set the following variants:

To verify the crucial roles played by our FCEM module and adaptive encoder module in the overall model, we set up several variants. Firstly, we chose the original diffusion model as the baseline, which is referred to as "*baseline*" in Table 2. To demonstrate the pivotal roles of the FCEM module and adaptive encoder module in the entire model, we removed these two modules separately from the complete model, which are referred to as "*w/o FCEM*" and "*w/o adaptive encoder*" in Table 2. Finally, we compared these variants with the full model "*Ours (full model*)" for comprehensive analysis, and some results are shown in Figure 12.



Result

GT

Figure 10: In order to simulate scenarios where there are no background objects available as reference, we only utilize the foreground encoder module during the foreground shadow mask prediction stage, hence not referring to background shadow information. Consequently, the inferred shadow shapes will deviate significantly from the ground truth.

Table 1: Results of quantitative comparison on our testing set. "↑" indicates the higher the better, and "↓" indicates the lower the better. The best results are marked in bold.

Method	RMSE \downarrow	SSIM ↑	fMSE ↓	fSSIM ↑
SGDiffusion [29]	8.591	0.825	875.882	0.809
ARShadowGAN [28]	9.164	0.817	942.154	0.816
ObjectStitch [46]	9.357	0.773	1145.116	0.773
DucoNet [48]	7.346	0.861	454.213	0.915
DIH-GAN [2]	6.145	0.847	567.311	0.894
Ours	5.582	0.917	367.919	0.937

We trained these variants using the same training data and quantitatively evaluated their impact on the test results. The evaluation results are presented in Table 2. From the table, we can observe that: after introducing guided supervision with FCEM, the model's quantitative performance has made significant strides, sufficiently demonstrating the strong guiding role of the FCEM module in capturing texture details of foreground objects. Moreover, with the inclusion of the adaptive encoder, the model's performance has also noticeably improved compared to the original diffusion model, confirming its guiding role in the harmonization generation of foreground objects.

With the simultaneous introduction of both the FCEM and adaptive encoder modules, our full model achieved the best performance, demonstrating the effectiveness of our approach. Additionally, incorporating the FCEM and adaptive encoder modules into the original diffusion model significantly improves performance.

 Table 2: Ablation study of FCEM and adaptive encoder. The best results are marked in bold.

Method	RMSE ↓	SSIM ↑	fMSE ↓	fSSIM ↑
baseline	11.192	0.687	543.417	0.675
w/o FCEM	7.264	0.746	398.542	0.785
w/o adaptive encoder	8.437	0.727	485.534	0.841
Ours	5.582	0.917	367.919	0.937

4.4 Limitations

Our CFDiffusion still has the following limitations: (1) As shown in Figure 10, for scenarios lacking background object references or involving complex situations where shadows are cast onto intersecting planes, our model struggles to generate shadows effectively. (2) Due to computational costs and processing speed limitations, our method is currently not applicable to real-time video lighting harmonization, which is also one of our future directions for improvement.

4.5 Conclusion and Future Work

In this paper, we proposed an image composite method based on the diffusion model, which focuses on harmonizing the illumination inconsistency between foreground objects and background, while generating realistic shadows. We employed an adaptive encoder to extract style features from the background to guide the

Conference acronym 'XX, June 03-05, 2018, Woodstock, NY

Anon.



Figure 11: The visual results of harmonization experiments on iHarmony4 [7]. It can be seen that our results are closest to the Ground Truth.

Table 3: Results of quantitative comparison on iHarmony4. "↑" indicates the higher the better, and "↓" indicates the lower the better. The best results are marked in bold.

Method	RMSE↓	SSIM ↑	fMSE ↓	fSSIM ↑
CDT-Net [6]	6.847	0.804	379.187	0.858
Harmonizer [23]	6.308	0.854	410.847	0.821
DucoNet [48]	6.152	0.876	365.236	0.915
Ours	5.582	0.917	367.919	0.937





Figure 13: The experimental results on DESOBAv2 dataset. The area enclosed by the red box is where the shadow is expected to be generated.

Figure 12: Ablation experiment results. From (a) to (f), they are respectively the composite images, w/o FCEM, w/o adaptive encoder, our complete method, and GT.

diffusion model in better harmonizing the foreground. Specifically, we introduced a FCEM module to further improve the ability to preserve details of foreground content. Finally, we have conducted

experiments on our proposed IH-SG dataset, as well as the popular DESOBAv2 dataset and iHarmony4 dataset, demonstrating that our method achieves significant improvements. In the future, we will expand CFDiffusion to adapt to real-time video lighting harmonization and shadow generation.

CFDiffusion: Controllable Foreground Relighting in Image Compositing via Diffusion Model

Conference acronym 'XX, June 03-05, 2018, Woodstock, NY

987

988

989

990

991

992

993

994

995

996

997

998

999

1000

1001

1002

1003

1004

1005

1006

1007

1008

1009

1010

1011

1012

1013

1014

1015

1016

1017

1018

1019

1020

1021

1022

1023

1024

1025

1026

1027

1028

1029

1030

1031

1032

1033

1034

1035

1036

1037

1038

1039

1040

1041

1042

1043 1044

929 **REFERENCES**

930

931

932

933

934

935

936

937

938

939

940

941

942

943

944

945

946

947

948

949

950

951

952

953

954

955

956

957

958

959

960

961

962

963

964

965

966

967

968

969

970

971

972

973

974

975

976

977

978

979

980

981

982

983

984

985

- Omri Avrahami, Dani Lischinski, and Ohad Fried. 2022. Blended Diffusion for Text-driven Editing of Natural Images. In 2022 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/cvpr52688.2022. 01767
- [2] Zhongyun Bao, Chengjiang Long, Gang Fu, Daquan Liu, Yuanzhen Li, Jiaming Wu, and Chunxia Xiao. [n. d.]. Deep Image-based Illumination Harmonization. ([n. d.]).
- [3] Andrew Brock, Jeff Donahue, and Karen Simonyan. 2018. Large Scale GAN Training for High Fidelity Natural Image Synthesis. International Conference on Learning Representations, International Conference on Learning Representations (Sep 2018).
- [4] Wenyan Cong, Junyan Cao, Li Niu, Chenglong Zhang, Xuesong Gao, Zhiwei Tang, and Liqing Zhang. 2021. Deep Image Harmonization by Bridging the Reality Gap. Cornell University - arXiv, Cornell University - arXiv (Mar 2021).
- [5] Wenyan Cong, Li Niu, Jianfu Zhang, Jing Liang, and Liqing Zhang. 2021. BargainNet: Background-guided domain translation for image harmonization. In 2021 IEEE International Conference on Multimedia and Expo (ICME). IEEE, 1–6.
- [6] Wenyan Cong, Xinhao Tao, Li Niu, Jing Liang, Xuesong Gao, Qihao Sun, and Liqing Zhang. 2022. High-Resolution Image Harmonization via Collaborative Dual Transformations. In 2022 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/cvpr52688.2022.01792
- [7] Wenyan Cong, Jianfu Zhang, Li Niu, Liu Liu, Zhixin Ling, Weiyuan Li, and Liqing Zhang. 2020. DoveNet: Deep Image Harmonization via Domain Verification. In 2020 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/cvpr42600.2020.00842
- [8] Xiaodong Cun and Chi-Man Pun. 2020. Improving the Harmony of the Composite Image by Spatial-Separated Attention Module. *IEEE Transactions on Image Processing* (Jan 2020), 4759–4771. https://doi.org/10.1109/tip.2020.2975979
- [9] JulianJorgeAndrade Guerreiro, Mitsuru Nakazawa, and Björn Stenger. [n. d.]. PCT-Net: Full Resolution Image Harmonization Using Pixel-Wise Color Transformations. ([n. d.]).
- [10] Zonghui Guo, Dongsheng Guo, Haiyong Zheng, Zhaorui Gu, Bing Zheng, and Junyu Dong. 2021. Image Harmonization With Transformer. International Conference on Computer Vision, International Conference on Computer Vision (Jan 2021).
- [11] Hao Guoqing, Satoshi Iizuka, and Kiichi Fukui. 2020. Image Harmonization with Attention-based Deep Feature Modulation. British Machine Vision Conference, British Machine Vision Conference (Jan 2020).
- [12] Roy Hachnochi, Mingrui Zhao, Nadav Orzech, Rinon Gal, Ali Mahdavi-Amiri, Daniel Cohen-Or, and AmitHaim Bermano. 2023. Cross-domain Compositing with Pretrained Diffusion Models. (Feb 2023).
- [13] Yucheng Hang, Bin Xia, Wenming Yang, and Qingmin Liao. 2022. SCS-Co: Self-Consistent Style Contrastive Learning for Image Harmonization. (Apr 2022).
- [14] Kaiming He, Xiangyu Zhang, Shaoqing Ren, and Jian Sun. 2015. Delving Deep into Rectifiers: Surpassing Human-Level Performance on ImageNet Classification. In 2015 IEEE International Conference on Computer Vision (ICCV). https://doi. org/10.1109/iccv.2015.123
- [15] Liu He, Yijuan Lu, John Corring, Dinei Florencio, and Cha Zhang. 2023. Diffusionbased Document Layout Generation. (Mar 2023).
- [16] Jonathan Ho, Ajay Jain, Pieter Abbeel, and UC Berkeley. [n. d.]. Denoising Diffusion Probabilistic Models. ([n. d.]).
- [17] Yan Hong, Li Niu, and Jianfu Zhang. 2022. Shadow Generation for Composite Image in Real-World Scenes. Proceedings of the AAAI Conference on Artificial Intelligence (Jul 2022), 914–922. https://doi.org/10.1609/aaai.v36i1.19974
- [18] Phillip Isola, Jun-Yan Zhu, Tinghui Zhou, and Alexei A. Efros. 2017. Image-to-Image Translation with Conditional Adversarial Networks. In 2017 IEEE Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10. 1109/cvpr.2017.632
- [19] Yifan Jiang, He Zhang, Jianming Zhang, Yilin Wang, Zhe Lin, Kalyan Sunkavalli, Simon Chen, Sohrab Amirghodsi, Sarah Kong, and Zhangyang Wang. 2021. SSH: A Self-Supervised Framework for Image Harmonization. In Proceedings of the IEEE/CVF International Conference on Computer Vision. 4832–4841.
- [20] Tero Karras, Samuli Laine, and Timo Aila. 2019. A Style-Based Generator Architecture for Generative Adversarial Networks. In 2019 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/cvpr. 2019.00453
- [21] Kevin Karsch, Kalyan Sunkavalli, Sunil Hadap, Nathan Carr, Hailin Jin, Rafael Fonte, Michael Sittig, and David Forsyth. 2014. Automatic Scene Inference for 3D Object Compositing. ACM Transactions on Graphics (May 2014), 1–15. https://doi.org/10.1145/2602146
- [22] Bahjat Kawar, Shiran Zada, Oran Lang, Omer Tov, Huiwen Chang, Tali Dekel, Inbar Mosseri, and Michal Irani. 2022. Imagic: Text-Based Real Image Editing with Diffusion Models. (Oct 2022).
- [23] Zhanghan Ke, Chunyi Sun, Lei Zhu, Ke Xu, and RynsonW.H. Lau. 2022. Harmonizer: Learning to Perform White-Box Image and Video Harmonization. (Jul 2022).

- [24] Eric Kee, James F. O'brien, and Hany Farid. 2014. Exposing Photo Manipulation from Shading and Shadows. ACM Transactions on Graphics (Sep 2014), 1–21. https://doi.org/10.1145/2629646
- [25] Gihyun Kwon and JongChul Ye. 2022. Diffusion-based Image Translation using Disentangled Style and Content Representation. (Sep 2022).
- [26] Jean-Francois Lalonde and Alexei A. Efros. 2007. Using Color Compatibility for Assessing Image Realism. In 2007 IEEE 11th International Conference on Computer Vision. https://doi.org/10.1109/iccv.2007.4409107
- [27] Jun Ling, Han Xue, Li Song, Rong Xie, and Xiao Gu. 2021. Region-aware Adaptive Instance Normalization for Image Harmonization. In 2021 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/ cvpr46437.2021.00924
- [28] Daquan Liu, Chengjiang Long, Hongpan Zhang, Hanning Yu, Xinzhi Dong, and Chunxia Xiao. 2020. ARShadowGAN: Shadow Generative Adversarial Network for Augmented Reality in Single Light Scenes. In 2020 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/ cvpr42600.2020.00816
- [29] Qingyang Liu, Junqi You, Jianting Wang, Xinhao Tao, Bo Zhang, and Li Niu. 2024. Shadow Generation for Composite Image Using Diffusion model. arXiv preprint arXiv:2403.15234 (2024).
- [30] Xihui Liu, Dong Huk Park, Samaneh Azadi, Gong Zhang, Arman Chopikyan, Yuxiao Hu, Humphrey Shi, Anna Rohrbach, and Trevor Darrell. 2023. More Control for Free! Image Synthesis with Semantic Diffusion Guidance. In 2023 IEEE/CVF Winter Conference on Applications of Computer Vision (WACV). https: //doi.org/10.1109/wacv56688.2023.00037
- [31] Lingxiao Lu, Jiangtong Li, Junyan Cao, Li Niu, and Liqing2023 Zhang. [n. d.]. Painterly Image Harmonization using Diffusion Model. ([n. d.]).
- [32] Andreas Lugmayr, Martin Danelljan, Andres Romero, Fisher Yu, Radu Timofte, and Luc Van Gool. 2022. RePaint: Inpainting using Denoising Diffusion Probabilistic Models. In 2022 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/cvpr52688.2022.01117
- [33] Shitong Luo and Wei Hu. 2021. Diffusion Probabilistic Models for 3D Point Cloud Generation. In 2021 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/cvpr46437.2021.00286
- [34] Chenlin Meng, Ya-Ling He, Song Yang, Jiaming Song, Jiajun Wu, Jun-Yan Zhu, and Stefano Ermon. 2021. SDEdit: Guided Image Synthesis and Editing with Stochastic Differential Equations. *Cornell University - arXiv,Cornell University - arXiv* (Aug 2021).
- [35] Chong Mou, Xintao Wang, Liangbin Xie, Jian Zhang, Zhongang Qi, Ying Shan, and Xiaohu Qie. 2023. T2I-Adapter: Learning Adapters to Dig out More Controllable Ability for Text-to-Image Diffusion Models. (Feb 2023).
- [36] Alex Nichol, Prafulla Dhariwal, Aditya Ramesh, Pranav Shyam, Pamela Mishkin, Bob McGrew, Ilya Sutskever, and Mark Chen. [n. d.]. GLIDE: Towards Photorealistic Image Generation and Editing with Text-Guided Diffusion Models. ([n. d.]).
- [37] Ben Poole, Ajay Jain, Jonathan Barron, Ben Mildenhall, Google Research, and UC Berkeley. [n.d.]. DREAMFUSION: TEXT-TO-3D USING 2D DIFFUSION. ([n.d.]).
- [38] Alec Radford, JongWook Kim, Chris Hallacy, A. Ramesh, Gabriel Goh, Sandhini Agarwal, Girish Sastry, Askell Amanda, Pamela Mishkin, Jack Clark, Gretchen Krueger, and Ilya Sutskever. 2021. Learning Transferable Visual Models From Natural Language Supervision. *Cornell University - arXiv,Cornell University arXiv* (Feb 2021).
- [39] Robin Rombach, Andreas Blattmann, Dominik Lorenz, Patrick Esser, and Bjorn Ommer. 2022. High-Resolution Image Synthesis with Latent Diffusion Models. In 2022 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/cvpr52688.2022.01042
- [40] Nataniel Ruiz, Yuanzhen Li, Varun Jampani, Yael Pritch, Michael Rubinstein, and Kfir Aberman. 2022. DreamBooth: Fine Tuning Text-to-Image Diffusion Models for Subject-Driven Generation. (Aug 2022).
- [41] Chitwan Saharia, William Chan, Huiwen Chang, Chris Lee, Jonathan Ho, Tim Salimans, David Fleet, and Mohammad Norouzi. 2022. Palette: Image-to-Image Diffusion Models. In Special Interest Group on Computer Graphics and Interactive Techniques Conference Proceedings. https://doi.org/10.1145/3528233.3530757
- [42] Chitwan Saharia, Jonathan Ho, William Chan, Tim Salimans, David J. Fleet, and Mohammad Norouzi. 2022. Image Super-Resolution Via Iterative Refinement. IEEE Transactions on Pattern Analysis and Machine Intelligence (Jan 2022), 1–14. https://doi.org/10.1109/tpami.2022.3204461
- [43] Mathew Salvaris, Danielle Dean, and Wee Hyong Tok. 2018. Generative Adversarial Networks. 187-208. https://doi.org/10.1007/978-1-4842-3679-6_8
- [44] Yichen Sheng, Jianming Zhang, and Bedrich Benes. 2021. SSN: Soft Shadow Network for Image Compositing. In 2021 IEEE/CVF Conference on Computer Vision and Pattern Recognition (CVPR). https://doi.org/10.1109/cvpr46437.2021.00436
 [45] Jiaming Song, Chenlin Meng, and Stefano Ermon. 2020. Denoising Diffusion
- [45] Jianning Song, Chemin Meng, and Sterano Ernon. 2020. Denoising Diricision Implicit Models. arXiv: Learning,arXiv: Learning (Oct 2020).
 [46] Yizhi Song, Zhifei Zhang, Zhe Lin, Scott Cohen, Brian Price, Jianming Zhang,
- [46] YIZhi Song, Zhifei Zhang, Zhe Lin, Scott Conen, Brian Price, Jianming Zhang, SooYe Kim, and Daniel Aliaga. 2022. ObjectStitch: Generative Object Compositing. (Dec 2022).

Anon.

- [47] Kalyan Sunkavalli, Micah K. Johnson, Wojciech Matusik, and Hanspeter Pfister. 2010. Multi-scale image harmonization. ACM Transactions on Graphics (Jul 2010), 1–10. https://doi.org/10.1145/1778765.1778862
- [48] Linfeng Tan, Jiangtong Li, Li Niu, and Liqing Zhang. 2023. Deep Image Harmonization in Dual Color Spaces. (Aug 2023).
- [49] Ashish Vaswani, Noam Shazeer, Niki Parmar, Jakob Uszkoreit, Llion Jones, AidanN. Gomez, Lukasz Kaiser, and Illia Polosukhin. 2017. Attention is All you Need. Neural Information Processing Systems, Neural Information Processing Systems (Jun 2017).
- [50] Xiaolong Wang, Ross Girshick, Abhinav Gupta, and Kaiming He. [n. d.]. Non-local Neural Networks. ([n. d.]).
- [51] Ben Xue, Shenghui Ran, Quan Chen, Rongfei Jia, Binqiang Zhao, and Xing
 Tang. 2022. DCCF: Deep Comprehensible Color Filter Learning Framework for

High-Resolution Image Harmonization. (Jul 2022).

- [52] Binxin Yang, Shuyang Gu, Bo Zhang, Ting Zhang, Xuejin Chen, Xiaoyan Sun, Dong Chen, and Fang Wen. 2022. Paint by Example: Exemplar-based Image Editing with Diffusion Models. (Nov 2022).
- [53] Bo Zhang, Yuxuan Duan, Jun Lan, Yan Hong, Huijia Zhu, Weiqiang Wang, and Li Niu. 2023. Controlcom: Controllable image composition using diffusion model. arXiv preprint arXiv:2308.10040 (2023).
- [54] Han Zhang, Ian Goodfellow, DimitrisN. Metaxas, and Augustus Odena. 2019. Self-Attention Generative Adversarial Networks. International Conference on Machine Learning, International Conference on Machine Learning (May 2019).
- [55] Shuyang Zhang, Runze Liang, and Miao Wang. 2019. ShadowGAN: Shadow synthesis for virtual objects with conditional adversarial networks. *Computational Visual Media* (Mar 2019), 105–115. https://doi.org/10.1007/s41095-019-0136-1